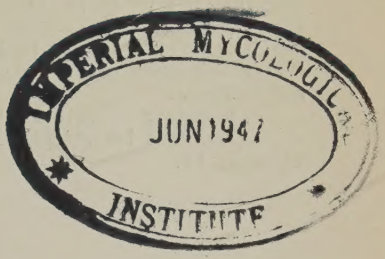
COMPTES RENDUS

(DOKLADY)

DE L'ACADÉMIE DES SCIENCES DE L'URSS

NOUVELLE SÉRIE

1945



VOLUME XLVIII

№ 8

ÉDITION DE L'ACADÉMIE DES SCIENCES DE L'URSS
MOSCOU

AVIS AUX AUTEURS

Les «Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS» paraissent tous les dix jours. Neuf numéros composent un volume pourvu d'une table des matières. Il y a quatre volumes par année.

Les «Comptes Rendus de l'Académie des Sciences de l'URSS» en de courtes communications précisent l'essentiel des travaux en préparation. Ces articles, ne dépassant pas en règle générale quatre pages, sont donnés en anglais ou en français. Ils embrassent les sciences mathématiques, physiques, naturelles et appliquées, mettant au jour les résultats de recherches scientifiques en cours ou qui viennent d'être achevées, et représentant ainsi les plus récentes données de l'investigation scientifique.

L'auteur a droit à 50 exemplaires de son article.

Adressez les manuscrits à la Rédaction des «Comptes Rendus», Moscou, Wolkhonka, 14.

20 Septembre

COMPTES RENDUS

(DOKLADY)

DE L'ACADÉMIE DES SCIENCES DE L'URSS

COMITÉ DE RÉDACTION

A. Arbusov, de l'Académie, **V. Chlopin**, de l'Académie, **A. Deborin**, de l'Académie, **A. Frumkin**, de l'Académie, **A. Kolmogoroff**, de l'Académie (vice rédacteur), **L. Orbeli**, de l'Académie (rédacteur en chef), **A. Richter**, de l'Académie, **S. Soboleff**, de l'Académie, **S. Vavilov**, de l'Académie (vice rédacteur)

NOUVELLE SÉRIE

13-me année

Paraissant tous les dix jours

1945

VOLUME XLVIII, № 8

TABLE DES MATIÈRES

MATHÉMATIQUES Pages

W. Kondrachov. Sur certaines propriétés des fonctions dans l'espace 535
A. Norden. The Affine Connectivity on the Surfaces of a Projective and Conformal Space 539
S. L. Soboleff, de l'Académie. Sur la presque périodicité des solutions de l'équation des ondes. I 542

MÉCANIQUE

P. K. Konakov. Experiment in the Application of the Theory of Similarity to Investigation of the Working Process in the Steam Engine of a Locomotive 546

HYDROMÉCANIQUE

A. M. Feinsilber. Integration of Boundary Layer Equations for a Compressible Fluid by the Method of Universal Variables 549

PHYSIQUE

E. Adirovich. Oscillation and Relaxation Processes in the Electron Plasma of the Discharge 551
M. G. Mescheryakov. On Absorption of Fast Neutrons by Heavy Nuclei . . 555

PHYSIQUE APPLIQUÉE

J. B. Friedman and **T. A. Volodina.** On the Strength and Ductility of Quenched and Tempered Steel 559
I. I. Kitaigorodsky. Electron Microscope and Investigation of the Structure of Ceramic Materials 563

CHIMIE PHYSIQUE

G. V. Akimov, Corresponding Member of the Academy, and **Z. L. Mintz.** Structure of Austenitic Heat-resistant Steel. Processes Involved in the Aging of Hardened EI-69 Steel 565
K. K. Andreev and **A. P. Bakeev.** On the Dependence of the Burning Velocity of Smokeless Powders upon the Pressure 567

CHIMIE BIOLOGIQUE

A. V. Blagoveshchensky and **A. J. Kologrivova.** Activation of Catalase by Biogenic Stimulants 571
M. V. Fedorov. Intensity of Fixation of Atmospheric Nitrogen by *Azotobacter* on Different Sources of Carbon 575
A. Passynsky and **V. Plaskeyev.** Diffusion and Molecular Weight of Lysozyme 579

GÉOLOGIE

E. V. Pavlovsky and **G. K. Semiguzov.** Quaternary Glaciation of the Sunnaghin Range (South Yakutia) 582
A. N. Solovkin. New Data on the Occurrence of Ores in the Intrusion of the Little Caucasus (Azerbaijan SSR) 585

MINÉRALOGIE

A. M. Kuznetzov and **D. P. Prochukhan.** On the Formation of Columnar Calcite 587
N. V. Renngarten. A Zeolite from the Mordenite Group in the Upper Cretaceous and Palaeogene Marine Deposits of the Eastern Slope of the Urals 591

PHYSIOLOGIE

T. I. Privolnev. Skin Respiration in *Carassius carassius* 594
A. A. Shutova. Phagocytic Activity of Leucocytes as Affected by Serums Containing Leucocytic Trephones 597

ENDOCRINOLOGIE

P. A. Wunder. On the Influence of Estrogens upon the Thyrotropic Reaction 602

MORPHOLOGIE EXPÉRIMENTALE

G. V. Lopashov. On some Simplest Processes of Organization of Eye and Brain Rudiment in Urodelan Embryos 605

PARASITOLOGIE

M. N. Nikolskaia. Evolution of *Eurytoma* Species *(Hymenoptera, Chalcididae)* in Connexion with the Geographical Distribution of their Food-plants of the subfam. *Prunoideae* 609

Traductions rédigées par D. Rakhmanov et T. Rogalina

Édition de l'Académie des Sciences de l'URSS
Moscou 1945

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

MATHÉMATIQUES

SUR CERTAINES PROPRIÉTÉS DES FONCTIONS DANS L'ESPACE

Par W. KONDRACHOV

(Présenté par S. L. Soboleff, de l'Académie, le 21. VI. 1945)

L_p^{ν} $(p \geqslant 1)$ est l'espace fonctionnel linéaire des fonctions $\varphi(x_1, \ldots, x_n)$ définies dans un domaine D de l'espace à n dimensions et dont les S-dérivées* jusqu'à l'ordre ν sont à p sommables ([1], [2]). Dans la présente Note D est un domaine borné sauf les cas que nous signalerons particulièrement.

1. Introduisons dans L_p^{ν} une métrique au moyen de la formule

$$\rho(\varphi, \psi) = \left[\overbrace{\int \cdots \int}^{n}_{D} \sum_{l=0}^{\nu} \sum_{\Sigma \alpha = l} \left| \frac{\partial^l (\varphi - \psi)}{\partial x_1^{\alpha_1} \ldots \partial x_n^{\alpha_n}} \right|^p dx_1, \ldots dx_n \right]^{\frac{1}{p}}$$

2. Indiquons quelques propriétés simples des espaces considérés:

a) L_p^{ν} est un espace complet;

b) si l'on a pour tous les α simultanément

$$\frac{\partial^{\alpha} \varphi_i}{\partial x_1^{\alpha_1} \ldots \partial x_n^{\alpha_n}} \xrightarrow{\text{faiblement}} \frac{\partial^{\alpha} \varphi}{\partial x_1^{\alpha_1} \ldots \partial x_n^{\alpha_n}}$$

au sens de l'espace $L_p\left(\sum \alpha_j = \alpha = 0, 1, \ldots, \nu\right)$, il en resulte que $\varphi_i \xrightarrow{\text{faiblement}} \varphi$ dans L_p^{ν} $(p > 1)$;

c) la sphère $\rho(0, \varphi) \leqslant A$ est faiblement compacte dans L_p^{ν} $(p > 1)$.

On désigne dans la suite par $\overline{L}_p^{\nu}$ la multiplicité linéaire de tous les éléments de L_p^{ν} qui sont des fonctions possédant des dérivées continues jusqu'au ν-ième ordre. On a

d) $\overline{L}_p^{\nu}$ est partout dense dans L_p^{ν}.

Pour les domaines D dont chaque point est accessible au moyen d'un secteur sphérique mobile dont la forme et la grandeur sont fixes, S. Soboleff a démontré le théorème suivant.

3. La sphère unitaire de L_p^{ν} est une partie de la sphère: 1) de chaque espace $L_1^{\nu - k}$ si $1 > 1/p > k/n$; 2) d'un espace $L_{q^{**}}^{\nu - k}$ quelconque, $q^{**} > 0$, si $\frac{1}{p} - \frac{k}{n}$

* On appelle S-dérivés les dérivés au sens de S. Soboleff [voir([1]), p. 483].

$1 > 1/p = k/n$; 3) de chaque espace $L_{q^*}^{\nu-k}$, où $q^* < q = 1/(1-k/n)$, si $p=1$; 4) d'un espace arbitraire $C^{\nu-k} = C^{\nu-[n/p]-1}$, si $1/p < k/n^*$; enfin, 5) de l'espace C pourvu que $n=k$, $p=1$.

4. Dans la présente Note on établit la proposition fondamentale suivante: L'ensemble $\rho(0, \varphi) \leqslant A$, borné dans L_p^ν, est compact dans les ensembles suivants qui le contiennent: 1) dans $L_{q^*}^{\nu-k}$, $q^* < q = 1/(1/p - k/n)$, si $1 > 1/p > k/n$; 2) dans $L_{q^{**}}^{\nu-k}$, $q^{**} > 0$, quand $1 > 1/p = k/n$; 3) dans $L_{q^*}^{\nu-k}$, $q^* < q = 1/(1-k/p)$, si $p=1$; 4) dans $C^{\nu-k} = C^{\nu-[n/p]-1}$, où $1/p < kn$ ([1], [3]).

Dans le cas où $p > 1$ les fonctions φ limites des suites φ_i que l'on peut extraire en vertu de la compacité indiquée, appartiennent à L_p^ν. La convergence des ν-ièmes dérivées des éléments φ_i vers les S-dérivées de la fonction φ est faible (au sens de L_p).

On peut modifier les énoncés des théorèmes **3** et **4** pour L_p^ν.

Considérons l'ensemble des fonctions φ vérifiant les conditions:

$$1) \qquad \overbrace{\int \dots \int}^{n-1}_{S_{n-1}} \left| \frac{\partial^l \varphi(x_1', \dots, x_n')}{\partial x_1^{\alpha_1} \dots \partial x_n^{\alpha_n}} \right| dS_{n-1} \leqslant A \quad (\textstyle\sum \alpha = l = 0, 1, \dots, \nu-1)$$

S_{n-1} est un morceau d'une multiplicité unie à $n-1$-dimensions, situé ou bien dans D, ou bien sur la frontière de D; $x_1', \dots, x_n'$ est un point intérieur à D, proche du point de S_{n-1}, qui lui correspond:

$$2) \qquad \overbrace{\int \dots \int}^{n}_{D} \sum_{\Sigma\alpha=\nu} \left| \frac{\partial^\nu \varphi}{\partial x_1^{\alpha_1} \dots \partial x_n^{\alpha_n}} \right|^p dx_1 \dots dx_n \leqslant A$$

L'ensemble ainsi extrait est borné dans L_p^ν.

5. Considérons l'espace L_p^k. Supposons que les fonctions $\varphi \subset L_p^k$, $\rho(0, \varphi) \leqslant A$, sont nulles hors du domaine D. Convenons de désigner dans la suite par C avec des indices inférieurs différents des constantes différentes, qui ne dépendent pas des fonctions considérées et des quantités h_1 et δ introduites. Posons encore $M+h = (x_1+h, \dots, x_n+h)$; $M = (x_1, \dots, x_n)$, $dx_1 \dots dx_n = dm$. Posons $\max |h_i| = \delta$ $(i = 1, 2, \dots, n)$. On a les évaluations suivantes.

a) Le domaine D est borné

$$\overbrace{\int \dots \int}^{n}_{D} |\varphi(M+h) - \varphi(M)|^{q^*} dm \leqslant$$

$$\leqslant C_{01} \begin{cases} \delta^{(1-q^*/q)}, & \dfrac{1}{q^*} > \dfrac{1}{q} = \dfrac{1}{p} - \dfrac{k}{n}, \quad p > 1, \quad \dfrac{1}{p} > \dfrac{k}{n} \\ \delta^{(1-q^*/q^{**})}, & q^{**} > 0 \text{ quelconque}, \quad \dfrac{1}{p} = \dfrac{k}{n} \\ & \dfrac{1}{p_1} - \dfrac{1}{p} = \dfrac{1}{q^*}\left(\dfrac{n-1}{n}\right), \quad p_1 \text{ donne } q^* \\ \delta^{(1-q^*/q_1^*)}, & q^* < q_1^* < \dfrac{1}{1-\dfrac{k}{n}}, \quad p = 1 \end{cases}$$

---

* Pour $n=2$ et en considérant une classe de domaines D plus restreinte, Reilich (Göttingen) a démontré la compacité de L_2' dans L_2 seulement.

b) Le domaine D est tout l'espace dont on peut enlever des morceaux finis de multiplicité assez unie de dimension inférieure à n.

$$1 < p \leqslant \frac{n}{k}$$

$$\overbrace{\int \cdots \int}^{n}_{D} |\varphi(M+h)-\varphi(M)|^{q^*} dm \leqslant C_{02} \begin{cases} \delta^{q^*} & \frac{1}{q^*} > \frac{1}{p} - \frac{k-1}{n} \\ \delta^{q^*} |\ln \delta|^{\frac{q^*(p-1)}{p}} & \frac{1}{q^*} = \frac{1}{p} - \frac{k-1}{n} \\ \delta^{q^*\left[k-n\left(\frac{1}{p}-\frac{1}{q}\right)\right]} & \frac{1}{q^*} < \frac{1}{p} - \frac{k-n}{n} \end{cases}$$

Si $p = \frac{n}{k}$, il faut appliquer les évaluations en prenant $p' < p$ de sorte que $1/p_1 - 1/p = \frac{1}{q^*}\left(\frac{n-1}{n}\right)$.

$$p = 1$$

$$\overbrace{\int \cdots \int}^{n}_{D} |\varphi(M+h)-\varphi(M)|^{q^*} dm \leqslant C_{03}\delta^{q^*(1-\varepsilon)}, \quad \frac{1}{q^*} > 1 - \frac{k+\varepsilon-1}{n}$$

c) $p > n/k$ ([2],[3]), $k = (m+n)/p$, $m > 0$.

$$|\varphi(M+h)-\varphi(M)|^{p} \leqslant C_{04} \begin{cases} \delta^{m} & m < p \\ \delta^{p} |\ln \delta|^{p-1} & m = p \\ \delta^{p} & m > p \end{cases}$$

Il est utile d'indiquer tout de suite brièvement le plan de la démonstration. On raisonne pour $\overline{L_p^k}$. On affirme que les évaluations précédentes sont vraies. Les inégalités 5a) et 5c), ainsi que les formules 1) —4) du § 3 permettent d'appliquer le théorème de Ritz (condition de la compacité dans L_p) et d'Arzela pour démontrer la proposition dans les domaines bornés. Ce qui est établi subsiste en tout cas dans chaque partie bornée des domaines non bornés que l'on considère. Les fonctions moyennes permettent de transporter le résultat dans L_p^ν. Dans cette direction on peut, en s'appuyant sur les propriétés des espaces (que l'espace est complet, *etc.*), utiliser le théorème de Fréchet.

Indiquons seulement l'idée de la démonstration des évaluations indiquées. Cette idée est basée sur l'application de l'identité de S. Soboleff ([2]), qui représente la valeur d'une fonction φ dans le sommet d'un certain secteur au moyen des valeurs des dérivées à l'intérieur d'un tel secteur; au cours de la démonstration il est commode d'utiliser le théorème de S. Soboleff sur les intégrales du type de potentiel ([1]).

Indiquons encore qu'en généralisant le théorème fondamental sur les intégrales du type de potentiel ([1]), sans perdre de vue le sens des raisonnements faits, on peut obtenir pour $\overline{L_p^\nu}$ des évaluations liées aux multiplicités à un nombre diverse de dimensions (ces évaluations complètent celles qui ont été publiées dans([2],[3])).

On peut aussi généraliser dans la direction de la présente Note les théorèmes de I. Petrovsky et K. Smirnoff (sur les fonctions également continues). Des relations analogues aux inégalités de Fridrichs permettent de lier l'énoncé de la proposition 4 avec les conditions concernant les espaces à $n-1$ - dimensions.

Quelques applications des résultats exposés seront considérés séparément.

Manuscrit reçu
le 21. VI. 1945.

LITTÉRATURE CITÉE

[1] S. Soboleff, Rec. Math., 4 (46), № 3 (1938). [2] S. Soboleff, C. R. Acad. Sci. URSS, I (X), 7 (84), 267 (1936). [3] W. Kondrachoff, C. R. Acad. Sci. URSS, XVIII, 236 (1938).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

MATHEMATICS

THE AFFINE CONNECTIVITY ON THE SURFACES OF A PROJECTIVE AND CONFORMAL SPACE

By A. NORDEN

(Communicated by A. N. Kolmogoroff, Member of the Academy, 13. X. 1944)

The surface X_m of m dimensions embedded in P_n* we shall call normalized, if with every of its points ** are connected two projective planes: 1) P_{n-m} containing this point and not incidental with the tangent plane T_m of the surface; 2) P_{m-1} belonging to T_m, but not containing the point of the surface.

We shall call these planes normalizing manifolds of the first and the second kind and denote them by I and II, respectively; the normalized surface we shall denote by A_m.

Introducing on A_m curvilinear coordinates u^i, $i = 1, 2, \ldots, m$, we define the plane II by the k points

$$y_i^\alpha = \partial_i x^\alpha - l_i x^\alpha \tag{1}$$

belonging to this plane.

When x^α is multiplied by the functional factor σ, the vector l_i must undergo a transformation according to the following law

$$\tilde{l}_i = l_i - \partial_i \lg \sigma \tag{2}$$

if the plane II remains invariant.

The plane I we define by $n-m+1$ independent points

$$x^\alpha, \quad \overset{s}{X}{}^\alpha, \qquad s = 1, 2, \ldots, n-m$$

The independence of the points x^α, y_i^α, $\overset{s}{X}{}^\alpha$ enables us to state the equations

$$\partial_j y_i^\alpha = A_{ij}^k y_k^\alpha + p_{ij} x^\alpha + \sum_{s=1}^{n-m} \overset{s}{B}_{ij} \overset{s}{X}_\alpha \tag{3}$$

where p_{ij} and $\overset{s}{B}_{ij}$ are tensors.

Alternating by the indices i and j and taking into account the equations (1), we find the symmetry of the expressions

$$G_{ij}^k = A_{ij}^k - l_j \delta_i^k \tag{4}$$

with respect to these indices.

---

* $B_n P_n$, P_m, *e.c.* we shall denote a projective space or a projective plane of the corresponding number of dimensions.

** Small latin characters with greek indices denote homogeneous coordinates. The greek index runs through the values from 1 to $n+1$.

The quantities G_{ij}^{k} do not depend on the choice of the points $\overset{s}{X}{}^{a}$, remain invariant when x^a is multiplied by a functional factor, and are transformed as the coefficients of affine connectivity under the transformation of the curvilinear coordinates.

Thus, on every A_m is defined a symmetrical affine connectivity, which we shall call the inner connectivity or the inner geometry of the given A_m.

To every pseudo-vector * v^i of the manifold of curvilinear coordinates corresponds a point of the plane II

$$v^a = v^i y_i^a \tag{5}$$

If the pseudo-vector is described as translated parallel in the inner conductivity under the condition

$$\delta v^i = \lambda v^i \tag{6}$$

then for such a translation it is necessary and sufficient that the point v undergo an infinitely small displacement in the P_{n-m+1} containing the plane I.

The inner connectivity is completely characterized by this fact. The geodesics of this connectivity are characterized by that their osculating P_{n-m+1} contains the manifold I, whence follows that the inner connectivity undergoes a projective transformation, when the planes II are replaced and the planes I are left unchanged.

If $m=n$, *i. e.* the manifold X_m coincides with P_n, then for its normalization it is sufficient to define a family of hyperplanes II which are in a continuous correspondence with the points of P_n. The geodesics of the inner geometry of A_n will be straight lines, and consequently, the inner geometry will be projectively plane. Any projectively plane geometry lends itself to this interpretation, which coincides with that proposed by Bortolotti (1).

If the correspondence between the points and the hyperplanes of A_n is one-to-one, then we may by means of a correlative transformation construct another projectively Euclidean geometry which will, generally speaking, be different from the first one.

We shall call these geometries inner geometries of the first and the second kind, respectively, defined by the given dual correspondence.

If the dual correspondence is reducible to polarity with respect to a hypersurface of the second class, then both inner geometries coincide and are geometries of the space of constant curvature, interpreted after Klein. If to every point of P_n the same hyperplane II is correlated, then the inner geometry of A_n will be affinely Euclidean with the unproper hyperplane II.

For the tangential coordinates of the tangent hyperplanes of the hypersurface A_{n-1} we may establish equations similar to (4), and determine on it, as above, the second inner connectivity with the coefficients Γ_{ij}^{k} which shall, generally speaking, be different from the G_{ij}^{k}. Thus, on a normalized hypersurface there are defined two inner connectivities of the first and the second kind, respectively.

The coefficients of these connectivities satisfy the relations

$$\partial_k B_{ij} - G_{ki}^{l} B_{lj} - \Gamma_{kj}^{l} B_{il} = \omega_k B_{ij} \tag{7}$$

where $B_{ij} = \underset{1}{B}_{ij}$ is a symmetrical tensor which by the equation

$$B_{ij} du^i du^j = 0 \tag{8}$$

* *i. e.* a vector defined up to a scalar factor.

determines the asymptotic cone of the hypersurface. The meaning of (7) is that the condition

$$B_{ij}v^i w^j = 0$$

for two pseudo-vectors to be adjoined remains valid, if each of them is parallelly translated in one of the two inner connectivities.

The inner geometry of any $A_m (m > 1)$ coincides with the inner geometry in the sense of Gauss, if we take for the absolute of Klein's space a certain hypersurface of the second class, and choose the plane I to be totally normal to the surface, whereas the plane II is considered as polar to the plane I with respect to the absolute.

As regards the hypersurface, its inner geometry of the second kind will be in this case also Riemannian with the linear element expressing the angle between the tangent hyperplanes in its adjacent points.

The inner connectivity of any $A_m (m>1)$ coincides with the connectivity induced by the pseudo-normal plane I [2], if all manifolds II belong to the unproper hyperplane of the affine space.

The inner geometry of the second kind of the hypersurface A_{n-1} will be in this case projectively plane and coincides with the inner geometry of the second kind on the unproper hyperplane, defined by the dual correspondence between the planes II and the unproper points of the straight lines I.

When this correspondence is polarity with respect to the absolute of the Euclidean space, then the geometry of the second kind of A_{n-1} is the geometry of its spherical representation.

The surface X_m in the space M_n of the conformal group of Moebius shall be called normalized and denoted by W_m, if in its every point is defined a sphere S of $n-m$ dimensions, passing through this point and totally normal to the surface.

Let us use the possibility to map the spheres of M_n into points of P_{n+1} so that the points of M_n are transformed into the points of a hyperquadric Q_n. X_m will then be transformed into a surface belonging to Q_n, and the sphere S into the section of Q_n by the plane P_{n-m+1}. Taking this plane for the manifold I and choosing for II its polar with respect to Q_n, we obtain a certain A_m in the space P_{n+1} as the representation of the given W_m. The inner geometry of A_m we shall call inner geometry corresponding to W_m of the space of Moebius. This geometry will be Weyl's geometry, and its angular metrics will coincide with the metrics induced on X_m by the other space M_n. If $m=n$, *i. e.* X_m coincides with M_n, then for its normalization it is sufficient to correlate to its every point another normalizing point.

Thus, every continuous point-to-point correspondence determines in M_n a certain geometry of Weyl, which will be conformably Euclidean. Conversely, every geometry of this type may be interpreted by means of the construction of a certain point-to-point correspondence in M_n.

If this correspondence is reducible to an inversion with respect to a hypersphere, then the inner geometry determined by it will be the geometry of a space of constant curvature interpreted by Poincaré's scheme.

Received
13. X. 1944.

REFERENCES

[1] E. Bortolotti, Annali di matematica pura ed applicata, IV, s. 11, 111 (1932).
Схоутен и Стройк, Введение в новые методы дифференц. геометрии, ч. 1, § 9, 1939.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

MATHÉMATIQUES

SUR LA PRESQUE PÉRIODICITÉ DES SOLUTIONS DE L'ÉQUATION DES ONDES. I

Par S. L. SOBOLEFF, de l'Académie

En cherchant la solution de l'équation des ondes

$$\Delta u - \frac{\partial^2 u}{\partial t^2} = 0 \tag{1}$$

dans un domaine Ω de variables indépendantes: $x_1, x_2, \ldots, x_n$, borné par l'hypersurface S et quand les conditions limites sont

$$\text{I)} \quad u|_S = 0$$

ou (2)

$$\text{II)} \quad \frac{\partial u}{\partial n}\Big|_S = 0$$

on emploie bien souvent la méthode de Fourier de la séparation des variables. On obtient de cette façon la solution générale du problème dans la forme suivante:

$$u(x_1, \ldots, x_n, t) = \sum_{j=1}^{\infty} a_j e^{i\lambda_j t} U_j(x_1, x_2, \ldots, x_n) \tag{3}$$

Dans l'espace H de Hilbert convenablement construit dont les éléments sont des fonctions des n variables $x_1, \ldots, x_n$ la solution (3) sera une fonction presque périodique de la variable t.

Pour $n = 1$ Muckenhoupt (1) a demontré, au moyen de considérations générales et sans recourir à la séparation des variables indépendantes, que la solution est une fonction presque périodique.

Nommons trajectoire de la solution l'ensemble de toutes les valeurs de $u(x_1, \ldots, x_n, t)$ dans H pour des valeurs différentes de t. S. Bochner (2) et J. v. Neumann (3) ont démontré que pour les équations d'une classe assez générale les solutions dont la trajectoire est compacte, sont presque périodiques.

Dans la présente Note nous donnons une démonstration complète de la presque-périodicité des solutions de l'équation (1), dans les conditions I) ou II), en faisant quelques hypothèses supplémentaires sur le contour S et les solutions examinées, hypothèses qui se rattachent à l'existence de certaines dérivées continues. Dans les Notes suivantes nous nous débarrassons de ces restrictions.

Supposons que l'espace H est constitué des paires des fonctions $u_0(x_1, \ldots, x_n)$, $u_1(x_1, \ldots, x_n)$ telles que u_0 est absolument continue par rapport à chaque x_i presque partout, tandis que $\partial u_0/\partial x_j$ et u_1 sont à carré sommable.

Nous traitons comme équivalent à zéro l'élément pour lequel $u_0 = \text{const}$, $u_1 = 0$, sauf, peut-être, un ensemble de points de mesure nulle.

Soient h et g deux éléments quelconques et soit h la paire u_0 et u_1, et g la paire v_0 et v_1. Le produit scalaire est déterminé par la formule

$$(h, g) = \overbrace{\int \cdots \int}^{n} \left[\sum_{j=1}^{n} \frac{\partial u_0}{\partial x_j} \frac{\partial \overline{v_0}}{\partial x_j} + u_1 \overline{v_1} \right] d\Omega \tag{4}$$

Soit $u(x_1, x_2, \ldots, x_n, t)$ — la solution de l'équation (1) dans les conditions I) ou II) qui a des dérivées continues du second ordre.

Construisons l'ensemble ordonné $h(t)$ d'éléments de H qui dépend d'un paramètre t:

$$\begin{aligned} \tilde{u}_0(x_1, x_2, \ldots, x_n; t) &= u(x_1, x_2, \ldots, x_n; t) \\ \tilde{u}_1(x_1, x_2, \ldots, x_n; t) &= \frac{\partial u}{\partial t}(x_1, \ldots, x_n; t) \end{aligned} \tag{5}$$

nous le nommons trajectoire de la solution. Nous dirons que la fonction $u(x_1, x_2, \ldots, x_n; t)$ engendre cette trajectoire. $h(t)$ et $g(t)$ étant deux trajectoires, on a la relation connue:

$$\frac{d}{dt}(h, g) = 0 \tag{6}$$

L'équation (6) est nommée loi de conservation de l'énergie. De (6) il suit

$$(h, g) = \text{const}, \quad (h, h) = \|h\|^2 = \text{const} \tag{7}$$

De cette façon chaque trajectoire est entièrement située sur une certaine sphère de l'espace H. En désignant $(h(t), h(t)) = E_1(u)$, nous aurons $E_1(u) = \text{const}$.

Une fonctionnelle bilinéaire $L(h, g)$ est nommée intégrale d'un certain ensemble de trajectoires de notre problème, si elle reste constante pour deux trajectoires de cet ensemble. Le produit scalaire (h, g) est alors une intégrale de l'ensemble M_1 des solutions deux fois dérivables.

Construisons la fonctionnelle:

$$\begin{aligned} Y_2(h, g) = \int_{\Omega} \cdots \int \left(\sum_{j=1}^{n} \sum_{k=1}^{n} \frac{\partial^2 u_0}{\partial x_j \partial x_k} \frac{\partial^2 \overline{v_0}}{\partial x_j \partial x_k} + \sum_{j=1}^{n} \frac{\partial u_1}{\partial x_j} \frac{\partial \overline{v_1}}{\partial x_j} \right) d\Omega + \\ + \int_{S} \cdots \int \left[\sum_{j=1}^{n-1} \frac{1}{R_j} \frac{\partial u_0}{\partial l_j} \frac{\partial \overline{v_0}}{\partial l_j} + \left(\sum_{j=1}^{n-1} \frac{1}{R_j} \right) \frac{\partial u_0}{\partial n} \frac{\partial \overline{v_0}}{\partial n} \right] dS \end{aligned} \tag{8}$$

On désigne ici par R_j les rayons de courbure des sections normales principales de l'hypersurface S, et par $l_1, l_2, \ldots, l_{n-1}$ les directions des tangentes à ces sections normales principales. Le domaine où $Y_2(h, g)$ existe est évident.

La fonctionnelle $Y_2(h(t), g(t))$ conserve une valeur constante pour deux trajectoires quelconques $h(t)$ et $g(t)$ qui appartiennent à l'ensemble M_2, pour lequel la fonction u qui engendre a les dérivées continues du troisième ordre à l'intérieur de Ω et les dérivées continues du second ordre jusqu'au contour S. Autrement dit, $Y_2(h, t)$ est une intégrale de l'ensemble M_2 des trajectoires.

Indiquons brièvement la démonstration. Nous avons:

$$Y_2(h(t_2),\ g(t_2)) - Y_2(h(t_1),\ g(t_1)) = \int_{t_1}^{t_2} \frac{dY_2(h(t'),\ g(t'))}{dt'}\, dt' \tag{9}$$

En calculant dY_2/dt et en transformant l'intégrale de volume en une intégrale superficielle nous obtenons:

$$Y_2(h(t_2),\ g(t_2)) - Y_2(h(t_1),\ g(t_1)) =$$

$$= \int_{t_1}^{t_2} \left[\int_S \cdots \int \left\{ \frac{d}{dt} \left[\left(\sum_{j=1}^{n-1} \frac{1}{R_j} \right) \frac{\partial u}{\partial n} \frac{\partial \overline{v}}{\partial n} + \sum_{j=1}^{n-1} \frac{1}{R_j} \frac{\partial u}{\partial l} \frac{\partial \overline{v}}{\partial l} \right] + \Phi \right\} dS \right] dt \tag{10}$$

où

$$\Phi = \sum_{j=1}^{n} \sum_{k=1}^{n} \left(\frac{\partial^2 u}{\partial x_j\, \partial t} \frac{\partial^2 \overline{v}}{\partial x_j\, \partial x_k} + \frac{\partial^2 u}{\partial x_j\, \partial x_k} \frac{\partial^2 \overline{v}}{\partial x_j\, \partial t} \right) \cos nx_k \tag{11}$$

On démontre la formule (10) en premier lieu dans le cas où les dérivées de u jusqu'au troisième ordre sont continues dans le domaine fermé. Cela étant, on peut remplacer la surface S par une surface intérieure S^*. En passant à la limite dans l'expression de $Y_2^*(h(t_2),\ g(t_2))$ $Y_2^*(h(t_1),\ g(t_1))$, nous obtenons (10). En différentiant les conditions limites et en se servant de (1) on peut démontrer que pour le problème I)

$$\Phi = -\frac{d}{dt} \left[\left(\sum_{j=1}^{n-1} \frac{1}{R_j} \right) \frac{\partial u}{\partial n} \frac{\partial \overline{v}}{\partial n} \right] \tag{12}$$

et pour le problème II)

$$\Phi = -\frac{d}{dt} \left[\sum_{j=1}^{n-1} \frac{1}{R_j} \frac{\partial u}{\partial l_j} \frac{\partial \overline{v}}{\partial l_j} \right] \tag{13}$$

Posons

$$Y_2(h(t),\ h(t)) = E_2(u)$$

Il est aisé de voir que dans l'égalité

$$E_2(u) = A \tag{14}$$

la valeur de A est complètement déterminée par les conditions initiales. Du fait que $E_2(u)$ reste borné pour chaque t on peut déduire la compacité de la trajectoire.

Démontrons en premier lieu qu'il s'ensuit:

$$\overbrace{\int \cdots \int}^{n}_{\Omega} \sum_{j=1}^{n} \sum_{k=1}^{n} \left| \frac{\partial^2 u}{\partial x_j\, \partial x_k} \right|^2 d\Omega \leqslant C \tag{15}$$

où C est une constante indépendante de t.

Dans le cas où S est suffisamment régulière, l'intégrale superficielle dans l'expression E_2 peut être transformée de la façon suivante:

$$\Psi = \int_S \cdots \int \left[\sum_{j=1}^{n-1} \frac{1}{R_j} \left| \frac{\partial u}{\partial l_j} \right|^2 + \left(\sum_{j=1}^{n-1} \frac{1}{R_j} \right) \left| \frac{\partial u}{\partial n} \right|^2 \right] dS =$$

$$= \int_\Omega \cdots \int \left[\sum_{j=1}^{n} \sum_{k=1}^{n} \sum_{s=1}^{n} G_{jks} \left(\frac{\partial u}{\partial x_k} \frac{\partial^2 \overline{u}}{\partial x_j\, \partial x_s} + \frac{\partial \overline{u}}{\partial x_k} \frac{\partial^2 u}{\partial x_j\, \partial x_s} \right) + \right.$$

$$+\sum_{j=1}^{n}\sum_{k=1}^{n} H_{jk}\left(\frac{\partial u}{\partial x_j}\frac{\partial \bar{u}}{\partial x_k}+\frac{\partial \bar{u}}{\partial x_j}\frac{\partial u}{\partial x_k}\right)\Big]\, d\Omega \tag{16}$$

où G_{jks} et H_{jk} sont des fonctions bornées

$$|G_{jks}|\leqslant L, \quad |H_{jk}|\leqslant L \tag{17}$$

En appliquant à (16) l'inégalité connue

$$|ab|\leqslant \frac{M}{2}|a|^2+\frac{1}{2M}|b|^2 \tag{18}$$

qui est valable pour n'importe quel M, nous obtenons l'évaluation

$$\psi\leqslant n^2L(1+M)E_1(u)+\frac{nL}{M}\int\cdots\int_{\Omega}\sum_{j=1}^{n}\sum_{k=1}^{n}\left|\frac{\partial^2 u}{\partial x_j\,\partial x_k}\right|^2 d\Omega \tag{19}$$

Si nous posons $M=2nL$, nous aurons:

$$\psi\geqslant -n^2L(1+2nL)E_1(u)-\frac{1}{2}\int\cdots\int_{\Omega}\sum_{j=1}^{n}\sum_{k=1}^{n}\left|\frac{\partial^2 u}{\partial x_j\,\partial x_k}\right|^2 d\Omega \tag{20}$$

En tenant compte de (9) on obtient

$$\frac{1}{2}\int\cdots\int_{\Omega}\sum_{j=1}^{n}\sum_{k=1}^{n}\left|\frac{\partial^2 u}{\partial x_j\,\partial x_k}\right|^2 d\Omega\leqslant E_2(u)+n^2L(1+2nL)E_1(u) \tag{21}$$

Des relations (21) et (14) on déduit immédiatement l'inégalité (15).

L'inégalité (15) suivant Kondrachov [4] entraîne la compacité des fonctions $\partial u/\partial x_j$ et $\partial u/\partial t$ dans L_2, d'où résulte immédiatement la compacité de $h(t)$ dans H, c. q. f. d.

Le théorème de Bochner permet de conclure que la solution examinée est presque périodique.

Manuscrit reçu
le 11. V. 1945.

LITTÉRATURE CITÉE

[1] C. F. Muckenhoupt, J. of Mathematics and Physics, Massachusets Institute of Technology, VIII, No. 3 (1929). [2] S. Bochner, Acta Math., **62** (1934). [3] S. Bochner and J. v. Neumann, Ann. of Math., **36** (1935). [4] V. I. Kondrachov, C. R. Acad. Sci. URSS, XLVIII, No. 8 (1945).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

MECHANICS

EXPERIMENT IN THE APPLICATION OF THE THEORY OF SIMILARITY TO INVESTIGATION OF THE WORKING PROCESS IN THE STEAM ENGINE OF A LOCOMOTIVE

By P. K. KONAKOV

(Communicated by S. P. Syromyatnikov, Member of the Academy, 7. XII. 1941)

The main features of the working process in the steam engine of a locomotive are the mean indicator steam pressure in the cylinder p_i and the steam consumption per piston stroke u_1. The ratio of p_i to the surplus pressure in the boiler p is termed the indicator coefficient of the locomotive engine, and is denoted by the symbol ξ. The quantities ξ and u_1 are functions of a great number of independent variables, including the main dimensions of the cylinder and the slide-valve, the initial parameters of the steam, the parameters of the working regime of the engine, the pressure of the exhaust steam, *etc.*

For a locomotive of a given type ξ and u_1 depend upon the degree of filling of the cylinder ε and the velocity of motion of the locomotive W. The graphs of the functions $\xi = \varphi(\varepsilon, W)$ and $u_1 = \psi(\varepsilon, W)$ represent families of curves with ε as their parameter.

During the past few years a large amount of experimental material has been accumulated in locomotive testing for different initial steam parameters and dimensions. Unfortunately, this valuable material has not as yet been generalized as it should. In this paper an attempt is made to apply the theory of similarity to generalize experimental data on locomotive engine testing.

The working process of a steam engine may be divided into the following periods according to the action of the steam in the cylinder: 1) the admission period, 2) the expansion period, 3) the exhaust period, 4) the compression period.

On the basis of the thermodynamics and kinematics of the motive and distributive mechanisms of a steam engine, we can construct a set of differential equations for these periods of steam action, which makes it possible to determine the state of the steam in the cylinder for any given boundary conditions.

Let us take a cylindrical system of coordinates and write out the differential equations for the admission period. During the time interval $d\tau$, du kg of steam with a heat content of i_a enter the steam cylinder whose volume is V. As a result of this, the heat content of the steam inside of the cylinder increases by the value $d(ui)$. The heat equivalent of the work done by the steam during the time interval under question equals $AVdp$, where p is the steam pressure in the cylinder. As the steam comes into

contact with the cylinder walls, part of its heat dQ is transferred to these walls.

On the basis of the first principle of thermodynamics we may write

$$dui_a = d(ui) + AVdp + dQ \tag{1}$$

For the quantity of steam du which enters the cylinder during the time $d\tau$ we may write the following expression:

$$du = d\left(\frac{V}{v}\right) = \mu \frac{\omega}{v'} \sqrt{2g \frac{k}{k-1} p_a v_a \left[1 - \left(\frac{p}{p_a}\right)^{\frac{k-1}{k}}\right]}\, d\tau \tag{2}$$

where p_a and v_a are pressure and specific volume of fresh steam in the valve-box; ω, the area of the steam window opening; μ, the consumption coefficient for the steam jet passing from the valve-box into the steam cylinder; v', the specific volume of the steam entering the cylinder.

To equations (1) and (2) should be added the equations of motion of the piston and the valve and the equation of state of the steam, which are not written here for lack of space. The initial and boundary conditions are formulated by equations which are also not given here. Equations (1) and (2), and the corresponding equations for the other steam action periods, give a number of characteristic relations. Analysis shows that of the relations thus obtained only one has any essential influence over the working process in a locomotive engine. This relation may be written as follows

$$\Pi_r = \frac{da_3 a_a^* \varepsilon}{D^2 S_p (1+s_0) n} \sqrt{1 - \left(\frac{p_e}{p_a}\right)^{\frac{k-1}{k}}} \tag{3}$$

where D is the diameter of the cylinder; S_p, the stroke of the piston; s_0, the clearance coefficient; d, the diameter of the cylindrical valve; a_3, the maximum linear opening of the steam window; a_a^*, the velocity of sound in steam with the parameters p_a and v_a; p_e, the pressure of the exhaust steam; n, the number of r. p. m. of the engine.

From relations that are not characteristic we select the following:

$$\bar{\xi} = \frac{p_i}{p_a} \tag{4}$$

and

$$\bar{u} = \frac{u_1 R T_a}{D^2 S_p (1+s_0) p_a} \tag{5}$$

where T_a is the absolute temperature of the fresh steam in the valve box.

On the basis of the theorems of Kirpichev (1) and Euckingham we can write:

$$\bar{\xi} = \varphi(\Pi_r) \tag{6}$$

$$\bar{u} = \psi(\Pi_r) \tag{7}$$

These functions have been used as a basis for treatment of the direct data obtained from tests of locomotives of the Эм, СОк and Сум series.

Figs. 1 and 2 show the graphs of the functions $\bar{\xi} = \varphi(\Pi_r)$ and $\bar{u} = \psi(\Pi_r)$ for these locomotives. A direct examination of these graphs suggests the conclusion that the experimental points fit in quite well along one curve, although the main dimensions of the locomotive engines tested, their

working regimes and the initial parameters of the working steam varied considerably.

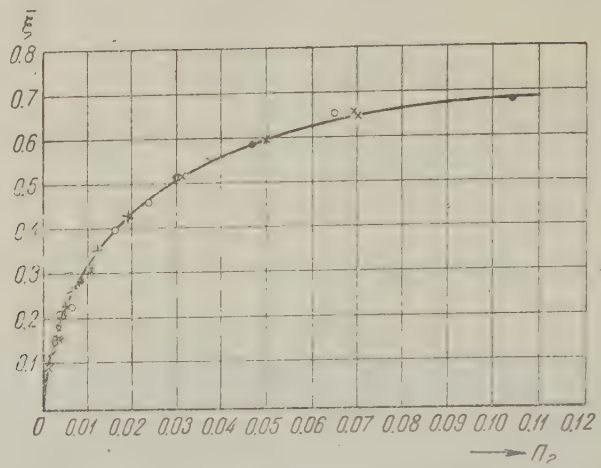


Fig. 1. ×, Эм locomotive; ●, СОк locomotive; ○, Сум locomotive

On the basis of the fundamental equations (6) and (7) a number of auxiliary equations can be constructed, which furnish quite an adequate picture of the operation of the locomotive engine.

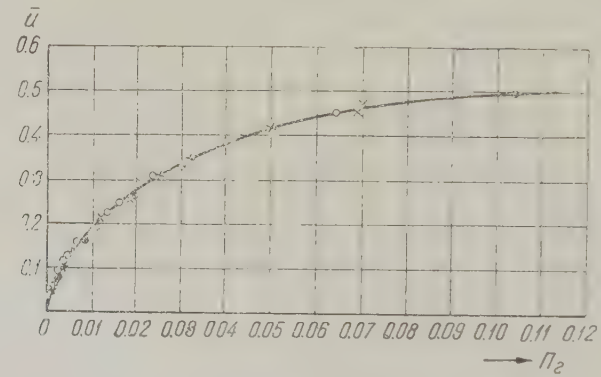


Fig. 2. × Эм locomotive; ●, СОк locomotive; ○, Сум locomotive

The relations established here make it possible to draw a number of important conclusions and can be used to great advantage in the modernization of current locomotives and in the designing of new ones.

Received
7. XII. 1944.

REFERENCES

[1] М. В. Кирпичев, Изв. Энергетич. ин-та АН СССР, 1 (1933).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

FLUID DYNAMICS

INTEGRATION OF BOUNDARY LAYER EQUATIONS FOR A COMPRESSIBLE FLUID BY THE METHOD OF UNIVERSAL VARIABLES

By A. M. FEINSILBER

(Communicated by N. E. Kochin, Member of the Academy, 7. IX. 1944)

In the following formulae u and v are the components of the velocity in the layer; ρ is density; p, pressure; θ, absolute temperature; τ, stress of friction; μ, the dynamic coefficient of viscosity; ν, the kinematic coefficient of viscosity; $\bar{u}$, $\bar{\mu}$, $\bar{\rho}$, *etc.* are the values of the respective quantities at the boundary of the layer; θ_0, ν_0, *etc.* are the values at the surface of the profile; Pr is the Prandtl number; Ba is the Barstow number; $\varkappa$ is the adiabatic index; $c = (\varkappa - 1)\,\mathrm{PrBa}^{-2}$.

In this notation the equations of the laminar boundary layer are

$$\rho u \frac{\partial u}{\partial x} + \rho v \frac{\partial u}{\partial y} = -\frac{dp}{dx} + \frac{\partial \tau}{\partial y} \tag{1}$$

$$\frac{\partial (\rho u)}{\partial x} + \frac{\partial (\rho v)}{\partial y} = 0 \tag{2}$$

$$c_p \left(\rho u \frac{\partial \theta}{\partial x} + \rho v \frac{\partial \theta}{\partial y} \right) = \frac{dp}{dx} u + \frac{c_p}{\mathrm{Pr}} \frac{\partial}{\partial y} \left(\mu \frac{\partial \theta}{\partial y} \right) + \mu \left(\frac{\partial u}{\partial y} \right)^2 \tag{3}$$

$$\tau = \mu \frac{\partial u}{\partial y} \tag{4}$$

$$\frac{\mu^2}{\bar{\mu}^2} = \frac{\theta}{\bar{\theta}} = \frac{\bar{\rho}}{\rho} \tag{5}$$

In integrating the system (1)--(5) we will use the method of universal variables, whose application to the case of incompressible fluid has been shown by the author in an earlier paper (1). Introduce x and u as independent variables, and

$$\tau_1 = \frac{\tau}{\bar{\mu}\,\bar{\rho}} \tag{6}$$

and

$$v_1 = \frac{\mu}{\bar{\mu}} \tag{7}$$

as functions.

In the new variables the system (1)–(5) reduces to the equations:

$$\tau_1^2 \frac{\partial^2 \tau_1}{\partial u^2} - \frac{\bar{u}\bar{u}'}{\bar{\mu}\bar{\rho}} \mu_1 \frac{\partial \tau_1}{\partial u} - \frac{u}{\bar{\mu}\bar{\rho}} \frac{1}{\mu_1} \frac{\partial \tau_1}{\partial x} + \left[\frac{u}{\bar{\mu}\bar{\rho}} \frac{\partial}{\partial x}\left(\frac{1}{\mu_1}\right) + \frac{\bar{u}\bar{u}'}{\bar{\mu}\bar{\rho}} \frac{\partial \mu_1}{\partial u} \right] \tau_1 = 0 \tag{8}$$

$$\tau_2 \left(\frac{\partial^2 \theta_1}{\partial u^2} + \frac{\mathrm{Pr}}{c_p \bar{\theta}} \right) + \frac{1-\mathrm{Pr}}{2} \frac{\partial \tau_2}{\partial u} \frac{\partial \theta_1}{\partial u} - \frac{2\mathrm{Pr}}{\bar{\mu}\bar{\rho}} u \frac{\partial \mu_1}{\partial x} - \frac{\mathrm{Pr}\,\bar{u}\bar{u}'}{\bar{\mu}\bar{\rho}} \mu_1 \frac{\partial \theta_1}{\partial u} = 0 \tag{9}$$

where $\tau_2 = \tau_1^2$; $\theta_1 = \mu_1^2$ (dimensionless temperature).

The boundary conditions of the system (8), (9) are

$$\mu_1 = \sqrt{\frac{\bar{\theta}_0}{\bar{\theta}}}\,; \quad \left(\tau_1 \frac{\partial \tau_1}{\partial u}\right) = -\frac{\bar{u}\bar{u}'}{\bar{\mu}\bar{\rho}} \mu_1 \qquad \text{for } u = 0 \tag{10}$$

$$\mu_1 = 1; \quad \tau_1 = 0 \qquad \text{for } u = \bar{u} \tag{11}$$

Making use for the confusor region of the expansions

$$\tau_1 = \sum_{i=0}^{\infty} a_i(x)\, u_1^i \tag{12}$$

$$\mu_1 = \sum_{i=0}^{\infty} b_i(x)\, u_1^i \tag{13}$$

we obtain from equations (8) and (9) and from boundary conditions (10), (11), a series of equations, of which only the first are written down here:

$$b_0 = \sqrt{\frac{\bar{\theta}_1}{\bar{\theta}}}\,; \quad \frac{a_0 a_1}{b_0} = -\frac{\bar{u}^2\bar{u}'}{\bar{\mu}\bar{\rho}}$$

$$\sum a_i(x) = 0; \quad \sum b_i(x) = 1$$

$$a_0^2 a_2 = \frac{\bar{u}^2 u'}{2\bar{\mu}\bar{\rho}} (b_0 a_1 - b_1 a_0)$$

$$a_0^2 \left(2b_0 b_2 + b_1^2 + \frac{c}{2} \right) = \frac{\bar{u}^2\bar{u}'}{\bar{\mu}\bar{\rho}} b_0^2 b_1 \ \textit{etc.}$$

Let us now find the velocity distribution. By virtue of (6), (12) and (13) we get

$$\frac{\bar{\rho}}{\bar{u}}\, y = \frac{b_0}{a_0} u_1 + \frac{1}{2}\left(\frac{b_1}{a_0} + \frac{a_1 b_0}{a_0^2} \right) u_1^2 + \ldots \tag{14}$$

The calculation of the separation will be discussed by the author in a subsequent paper.

Received
7. IX. 1944.

REFERENCES

[1] A. M. Feinsilber, C. R. Acad. Sci. URSS, XLVIII, No. 7 (1945).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

PHYSICS

OSCILLATION AND RELAXATION PROCESSES IN THE ELECTRON PLASMA OF THE DISCHARGE

By E. ADIROVICH

(Communicated by S. I. Vavilov, Member of the Academy, 25. X. 1944)

1. Langmuir([1]) and Tonks and Langmuir([2]) have shown that local (not travelling) oscillations with a frequency $\omega_L = \sqrt{4\pi ne^2/m}$ (Langmuir frequency), where m and n are mass and density of the electrons, are possible in the plasma. Linder([3]), upon taking the pressure of electronic gas into account, alongside the electrostatic forces, has found that longitudinal waves of electronic density can propagate in the plasma at frequencies higher than that of Langmuir. Vlassov ([4],[5]) has evolved general methods in the theory of oscillation processes in the plasma, in which the use of the kinetic equation is combined with a macroscopic description of the interaction between volume charges. However, owing to a mathematical error made in solving the equation of dispersion, Vlassov's results are only true for a branch of local oscillations with frequencies lower than the Langmuir frequency.

The theory put forward in the present paper proceeds upon a scheme which takes account of only «distant» forces, as proposed by Vlassov, but is complemented by the suggestion to compute the integral in the equation of dispersion in the sense of its Cauchy's principal value. It appears in this theory that density waves propagating with higher as well as lower frequencies compared to that of Langmuir are in existence in the plasma.

Within the scope of his theory are also the relaxation processes which correspond to the retardation of the beams by the plasma.

2. The fundamental equation

$$\frac{\partial \Phi}{\partial t} + u\frac{\partial \Phi}{\partial x} - \frac{4\pi ne^2}{k_0 t}\left(\frac{m}{2\pi k_0 T}\right)^{3/2} ue^{\frac{-m(u^2+v^2+w^2)}{2k_0T}} \times \times \int_{-l}^{x} dx \int_{-\infty}^{+\infty} \Phi du\, dv\, dw = 0 \qquad (1)$$

for the «disturbance» of the function of electron distribution

$$\Phi(x, u, v, w, t) = f(x, u, v, w, t) - n\left(\frac{m}{2\pi k_0 T}\right)^{3/2} e^{-\frac{m(u^2+v^2+w^2)}{2k_0T}} \qquad (2)$$

is obtained from the kinetic equation on the following assumptions: 1) the problem is one-dimensional, 2) the stationary distribution of electrons in the plasma is a Maxwellian distribution, 3) the disturbing influences on the

plasma are small and short-lived, 4) collisions are neglected. The solution of this equation is a superposition of expressions of the form

$$\Phi(x, u, v, w, t) = A \frac{ue^{-\frac{m(u^2+v^2+w^2)}{2k_0T}}}{u - k/q} e^{qx-kt} \tag{3}$$

the quantities k and q being connected by the equation of dispersion

$$\int_{-\infty}^{+\infty} \frac{ue^{-\frac{mu^2}{2k_0T}}du}{u - k/q} = \frac{k_0T}{2ne^2}\sqrt{\frac{k_0T}{2\pi m}}\, q^2 \tag{4}$$

This equation can be transformed into

$$\int_{-\infty}^{+\infty} \frac{xe^{-x^2/2}dx}{x - \tilde{k}/\tilde{q}} = \sqrt{2\pi}\,\tilde{q}^2 \tag{4'}$$

where

$$\tilde{k} = k\sqrt{\frac{m}{4\pi ne^2}} = \frac{k}{\omega_L};\quad \tilde{q} = q\sqrt{\frac{k_0T}{4\pi ne^2}} = \frac{q}{q_D} \tag{5}$$

If $\operatorname{Im}\tilde{k}/\tilde{q} \neq 0$, the integral in (4') is the limit for $b \to \infty$ of the corresponding integral of Cauchy, taken from $-b$ to $+b$ (non-singular case). If $\operatorname{Im}\tilde{k}/\tilde{q} = 0$, this integral is to be computed as the limit for $b \to \infty$ of the principal value of the corresponding integral of Cauchy type (singular case).

3. Let us consider the function

$$\psi(\tilde{v}, \beta) = \int_{-\infty}^{+\infty} \frac{e^{-\beta(x^2 - v^2)}}{x - \tilde{v}}\, dx \tag{6}$$

Differentiating with respect to β, we obtain the equation

$$\frac{\partial\psi}{\partial\beta} = -\tilde{v}\sqrt{\frac{\pi}{\beta}}e^{\beta\tilde{v}^2} \tag{7}$$

whence

$$\psi(\tilde{v}, \beta) = -\tilde{v}\sqrt{\pi}\int_0^{\beta} \frac{e^{\tilde{v}^2\xi}d\xi}{\sqrt{\xi}} + \int_{-\infty}^{+\infty} \frac{dx}{x - \tilde{v}} \tag{8}$$

Denoting $\tilde{k}/\tilde{q} = \tilde{v}$, expressing the integral in (4') through $\psi(\tilde{v}, {}^1/_2)$, substituting $\tilde{v}^2\xi = \eta^2$ in the first integral in (8) and computing $\int_{-\infty}^{+\infty} \frac{dx}{x - \tilde{v}} = \mp\pi i$, where the signs + and − correspond, respectively, to $\tilde{v}$ lying in the upper and lower half-plane of the plane of complex numbers, we at last obtain

$$\int_{-\infty}^{+\infty} \frac{xe^{-x^2/2}dx}{x - \tilde{v}} = \sqrt{2\pi} - 2\sqrt{\pi}\,\tilde{v}e^{-\tilde{v}^2/2}\int_0^{\tilde{v}^2/\sqrt{2}} e^{\eta^2}d\eta \mp \pi i\tilde{v}e^{-\tilde{v}^2/2} \tag{9}$$

With the plus and the minus sign before the last term the relation (9) describes $2\pi i\psi_i(\tilde{v})$ and $2\pi i\psi_e(\tilde{v})$, respectively, where $\psi_i(\tilde{v})$ and $\psi_e(\tilde{v})$ are func-

tions defined by the integral (4′) in the upper and lower half-plane of the plane of complex numbers. Cauchy's principal value of the integral for $\tilde{v}$ real is

$$\text{V.p.} \int_{-\infty}^{+\infty} \frac{xe^{-x^2/2}\,dx}{x-\tilde{v}} = \frac{2\pi i\varphi_i(\tilde{v}) + 2\pi i\varphi_e(\tilde{v})}{2} = \sqrt{2\pi} - 2\sqrt{\pi}\,\tilde{v}e^{-\tilde{v}^2/2} \int_{\tilde{v}/\sqrt{2}}^{\infty} e^{\eta^2}\,d\eta \quad (10)$$

The solutions of the equation of dispersion for the non-singular case are represented graphically in Figs. 1 and 2. In Tables 1, 2 and 3 approximate analytical expressions of these curves in the neighbourhood of certain points are given.

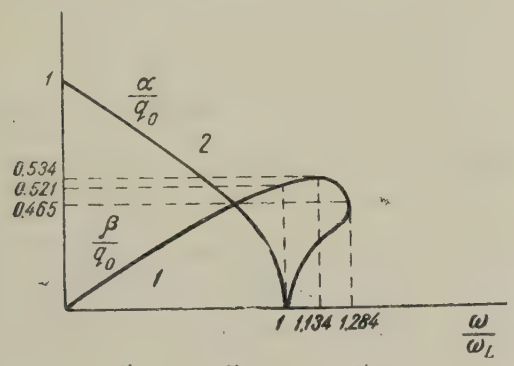


Fig. 1. $\tilde{q}_1 = \alpha/q_D$; $\tilde{q}_2 = i\beta/q_D$; $\tilde{k} = i\omega/\omega_L$

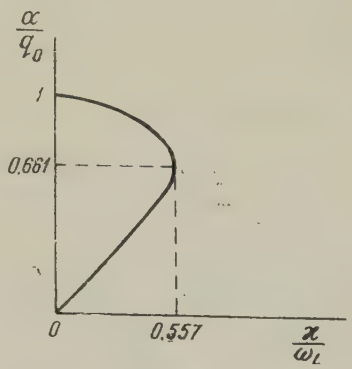


Fig. 2. Dependence of α/q_D from x/ω_L; $\tilde{q} = \alpha/q_D$; $\tilde{k} = x/\omega_L$

Together, the expression (3), the curves in Figs. 1 and 2, and the method of computing the approximate expressions for these curves in elementary functions constitute a complete solution of the problem of oscillation ($k = i\omega$) and relaxation ($k = x$) processes in the plasma of a discharge ([6]). Macroscopically these processes are characterized by the surplus density of electrons

$$\Delta n(x,t) = \int_{-\infty}^{+\infty} \Phi(x, u, v, w, t)\,du\,dv\,dw = Cq^2 e^{qx-kt} \quad (11)$$

and by the flow of electrons

$$S(x,t) = \int_{-\infty}^{+\infty} u\,\Phi(x, u, v, w, t)\,du\,dv\,qw = Ckqe^{qx-kt} \quad (12)$$

Table 1

Approximate Expressions for the β(ω) Curve (Fig. 1)

At $\frac{\omega}{\omega_L} \to$	$\frac{\beta}{q_D} \cong$		$\frac{\varphi}{\sqrt{k_0 l/m}} \to$
0	$0.765\,\frac{\omega}{\omega_L} - 0.448\,\frac{\omega^3}{\omega_L^3}$		1.307 ∓ 0.003
1	$0.521 + 0.174\left(\frac{\omega}{\omega_L} - 1\right)$	(upper branch)	1.916 ∓ 0.007
1	$\sqrt{\frac{\omega^2}{\omega_L} - 1}$	(lower branch)	∞
1.134	$0.534 - 0.936\left(\frac{\omega}{\omega_L} - 1.134\right)^2$		2.124 ∓ 0.003
1.284	$0.465 \mp 0.411\sqrt{1.284 - \frac{\omega}{\omega_L}}$		2.76 ∓ 0.01

Table 2

Approximate Expressions for the $\alpha(\omega)$ Curve (Fig.1)

At $\frac{\omega}{\omega_L} \longrightarrow$	$\frac{\alpha}{q_D} \cong$
0	$1 - \frac{1}{2}\sqrt{\frac{\pi}{2}}\,\frac{\omega}{\omega_L}$
1	$\frac{1}{\sqrt{3}}\,\frac{\omega}{\omega_L}\sqrt{1-\frac{\omega}{\omega_L}}$

Table 3

Approximate Expressions for the $\alpha(x)$ Curve (Fig.2)

At $\frac{x}{\omega_L} \rightarrow$	$\frac{\alpha}{q_D} \cong$		$\frac{\varphi}{\sqrt{k_0 t/m}} \rightarrow$
0	$1 - \frac{1}{2}\,\frac{x^2}{\omega_L^2}$	(upper branch)	0
0	$0.765\,\frac{x}{\omega_L} + 0.448\,\frac{x^3}{\omega_L^3}$	(lower branch)	1.307 ∓ 0.003
0.557	$0\,661 \mp 0.885\sqrt{0.557 - \frac{x}{\omega_L}}$		0.8422 ∓ 0.0007

4. What has been considered by Vlassov ([4],[5]) corresponds to a singular case of the dispersion equation, since for harmonic waves the ratio $\tilde{k}/\tilde{q}$ (ω^*/k^* in his notation) is a real number. Vlassov approximates the integral $\int_{-\infty}^{+\infty} \frac{xe^{-x^2/2}dx}{x-\omega^*/k^*}$ to the first terms of a series obtained by binomial expansion of the denominator of the integrand, on the assumption that $|x| < \omega^*/k^*$. This means that the above integral is taken in the limits from $-c$ to $+c$, where $c < \omega^*/k^*$. Now the equation of dispersion in ([4]), formula (39), contains an integral with infinite limits. Neglected, therefore, by Vlassov are not only the higher powers of the ratio ck^*/ω^*, but also the integrals $\int_{-\infty}^{-c} \frac{xe^{-x^2/2}dx}{x-\omega^*/k^*}$ and $\int_{c}^{\infty} \frac{xe^{-x^2/2}dx}{x-\omega^*/k^*}$, one of which must necessarily diverge for $\mathrm{Im}\,(\omega^*/k^*) = 0$. On this account Vlassov's formula for dispersion ([4], formula (50)) does not coincide with Linder's ([3], formula (7)), and his results are only valid for local oscillations with frequencies $\omega < \omega_L$, when ω^*/k^* is an imaginary number, and the residual term of the expansion contains no singularity in the denominator of the function under the integral sign.

Received
25. X. 1944.

REFERENCES

[1] I. Langmuir, Proc. Nat. Acad. Sci., 14, No. 8, 627 (1928). [2] L. Tonks and I. Langmuir, Phys. Rev., 33, No. 2, 195 (1929). [3] E. Linder, Phys. Rev., 49, No. 10, 753 (1936). [4] А. А. Власов. ЖЭТФ, 8, № 3, 291 (1938). [5] А. А. Власов, Thesis, НИИФ МГУ, 1942. [6] Э. И. Адирович, Thesis, НИИФ МГУ, 1944.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

PHYSICS

ON ABSORPTION OF FAST NEUTRONS BY HEAVY NUCLEI

By M. G. MESCHERYAKOV

(Communicated by V. G. Chlopin. Member of the Academy, 6. XI. 1944)

For fast neutrons the probability of being captured by heavy nuclei is extremely small in comparison with the competing process of inelastic scattering. Thus, for neutrons with energy $E > 1$ MeV and sufficiently heavy ($A > 50$) nuclei the cross-section for an (h, γ) process was obtained by Weisskopf and Ewing (1) in the form

$$\sigma(h, \gamma) = \pi R^2 \xi(E) \frac{\Gamma_\gamma}{\sum_\alpha \Gamma_\alpha},$$

where πR^2 is the geometrical section of the nucleus, $\xi(E)$ is the sticking probability; $\xi(E) \sim 1$ for $E > 1$ MeV. The factor $\Gamma_\gamma / \sum_\alpha \Gamma_\alpha$ represents the relative probability of a compound nucleus being destroyed by emission of a gamma-quantum: Γ_γ is partial width of level for such emission. The sum is to be taken over all partial widths of the particles that can possibly be emitted by a compound nucleus. By the above authors, $\sigma(h, \gamma)$ for heavy nuclei is estimated as $\sim 10^{-26}$ cm^2, if $E = 1$ MeV. Statistical models of the compound nucleus lead — as might well be expected, indeed, — to the conclusion that the factor $\Gamma_\gamma / \sum_\alpha \Gamma_\alpha$ is a smooth function of the excitation energy and atomic weight. From this and from simple considerations on the stability of nuclei it may be expected that for fast neutrons and heavy nuclei the probability of an (h, γ) process should increase monotonously with the atomic weight.

To prove how far these general considerations are true for neutrons from a (d, d)-reaction, was the object of the experiments described below, in which relative values of the cross-section for the (h, γ) process were determined for 19 elements. The measurements were made for those nuclei which by getting neutrons stuck to them form a familiar radioactive isotope.

Determining σ (h, γ) from the value of neutron-induced radioactivity rather than from the total weakening of the beam of neutrons, we leave processes of elastic and inelastic scattering of neutrons out of consideration. This method, however, has two drawbacks: 1) absorption of β-electrons in the thickness of the target and in the walls of the measuring apparatus, which compells us to measure the layer of semi-absorption of β-electrons in the material of targets, and 2) possibility of formation of isomeric activities with very short or very long periods. Under these conditions the value obtained for the cross-section can only be an underestimate.

The targets of the elements under investigation (Kahlbaum) were bombarded with the aid of a cyclotron by neutrons from a (d, d)-reaction. The maximum energy of the neutrons—which were emitted at an angle of 90° to the direction of the

ionic beam—was ~2. 70 MeV (energy of deutons, 1 MeV). Because of the presence of large masses of iron about the target (poles of the cyclotron's magnet), the spectrum of the neutrons showed diffusion toward the side of smaller energies. An inelastically scattered neutron has a mean energy approximately equal to the square root of the energy of the incident neutron(2), therefore secondary neutrons should have an energy of the order of 1—1.5 MeV. At the site of bombardment the intensity of the source of neutrons was equivalent to 20 Curie (Rn + Be). The neutrons, owing to the broadness of their energy spectrum, could be captured to many nuclear levels, whereby the influence of resonance and other features arising from the properties of individual states were eliminated. In experiments with fast neutrons the results can be greatly distorted by a slight admixture of slow neutrons, scattered by the surrounding objects containing hydrogen.

Therefore care was taken before starting the bombardment experiment that there should be no substances containing hydrogen in the vicinity of the cyclotron chamber; the targets were always surrounded by cadmium (0.5 g/cm^2) and boron (1.5 g/cm^2 in the form of B_4C). It was proved by careful control tests with rhodium and manganese that a boron screen causes no decrease of activity. Nor was the activity affected when a silver target was surrounded by a silver layer, or an iodine target by an iodine layer. All of this testified to the absence of slow (resonance) neutrons in the neighbourhood of the chamber.

Bombardment lasted 20 minutes in the case of V, Cu (5-min. period), Rh and Ag; 2.9 hours in the case of Mn, Br, I, Ba and Dy. The rest were elements of long period and were continuously bombarded for 36.3 hours. Control measurements showed that during long hours of continuous work of the cyclotron the intensity of neutronic emission fluctuated within 10—15 per cent. The activity of the targets, all of which had the same dimensions (3×5 cm^2), was measured by Geiger counters with an accuracy of ±3 per cent. Generally, the background of the counters did not surpass 5 per cent of the activities measured.

The cross-section for an (h, γ) process was computed from the formula

$$\sigma(h, \gamma) = \alpha \frac{NA}{\delta}$$

where N is the activity of a target which is thick relative to the absorption of β-electrons; A is atomic weight; δ is the layer of the half-absorption of β-electrons in the material of the target. The values of δ for the elements investigated were found by experiment, and are given here in the third column of the table.

As the geometric conditions for bombardment and for the measurement of activity were always the same, the coefficient α has one value for all targets. In measuring N corrections were introduced to allow for: 1) the finite length of the time of bombardment, 2) the decay between end of bombardment and beginning of measurement, 3) abundance of the initial isotope, 4) absorption of electrons by the walls (thickness 0.1 mm Al) of the counter. For Br, in the form of $PbBr_2$, and for rare earths, in the form of oxides, the presence of alien atoms in the material was also taken into account. The results of the experiments are listed in the fourth column of the table. As it was difficult in these experiments to estimate the exact number of the neutrons, the cross-sections for the (h, γ) process were expressed in a relative scale, the value of δ (h, γ) for vanadium being put at unity. In view of the many corrections, the error of the results should lie in the neighbourhood of 20 per cent. In the fifth and sixth column of the table are given by way of comparison, on the same relative scale, experimental data by Halban and Kowarski (3) for neutrons with an energy of 0.220±0.04 MeV (ThC″γ—D), and by Griffiths (4) for neutrons with energies 0.04 and 0.45 MeV (Raγ—Be).

Comparing their results with ours one may see that for neutrons with energies ranging from 0.04 to ~2 MeV the probability of an (h, γ) process shows a similar law of variation with atomic weight. Toward the end of the periodic table the value of the cross-section for the (h, γ) process shows a tendency to increase, which is in agreement with the theoretical conclusions. Yet this general tendency is often broken. For certain nuclei with odd Z and A (La, Pr, Bi) distributed without apparent order in the periodical table, the probability of an (h, γ) process is

Relative Scale of Cross-section Values for (h, γ) Process

Isotope under bombardment	Period	δ, g/cm^2	$\sigma(h,\gamma)$ (d, d)	σ (h,γ) (ThC″γ—D)	σ (h,γ) (Raγ—Be)
V^{51}	3.75 min.	0.130	1	1	1
Mn^{55}	2.5 hours	0.130	2	1	2.1
Ni		—	—	< 1	—
Cu^{63}	12.8 hours	0.03	1.3	—	1.7
Cu^{65}	5 min.	1.14	2	~ 1	1
Zn^{68}	13.8 hours	—	< 0.04	—	—
Zn^{68}	1 hour	—	< 0.04	—	—
As^{75}	26.8 hours	0.14	10.4	9.2	15.2
Se		—	—	< 0.3	—
Br^{79}	18 min.	0.08	25.8	17.8	23.3
Br^{79}	4.4 hours	0.08	12.6		12.9
Br^{81}	34 »	—	—	< 3.3	—
Y^{89}	60 »	0.078	—	—	0.9
Rh^{103}	44 sec	0.096	20	—	—
Rh^{103}	4.2 min.	0.096	5.2	—	—
Pd^{108}		—	—	27.8	—
Ag^{107}	2.3 min.	0.063	39.4	33.3	47.6
Ag^{109}	24 sec.	0.112	55.8	51.1	—
In^{115}	54 min.	0.038	—	—	22.4
Sb^{121}	2.8 days	0.073	25.8	12.2	30
Te		—	—	< 1	—
I^{127}	25 min.	0.076	19.6	19.4	32.9
Ba^{138}	86 »	—	< 0.1	0.2	—
La^{139}	42 hours	0.06	3	—	2
Pr^{141}	19 »	0.09	6.8	—	5.2
$Eu^{151,153}$	6 »	0.072	—	—	45.2
Tb^{159}	2.5 »	0.041	—	—	1
Tb^{159}	3 months	0.04	—	—	85.7
Dy^{164}	2.5 hours	0.035	13.8	< 7.8	—
Lu^{175}	3.5 »	0.036	—	—	81
$W^{184,186}$	23 »	0.047	13.2	2.9	—
$Ir^{191,193}$	19 »	0.07	—	—	62 or 95.2
Au^{197}	2.7 days	0.03	49.8	25.5	62
$Te^{203,205}$	2.5 min.	0.06	—	~ 1	6.2 or 2.9
Bi^{209}	5 days	—	< 0.04	—	—

extremely small. In some cases, however, there may be no real disturbance in the monotonous increase of δ (h, γ) with atomic weight, but merely one produced by isomery. For one of the silver isotopes, for example, the computed cross-section is less than the true one, since in silver bombarded by neutrons, an isomeric activity with a period of $\sim$225 days is observed, alongside of activities with 2.3 min. and 24 sec. periods. Subjecting La and Pr, which have low δ (h, γ) cross-sections, to prolonged bombardment, Griffiths found no indication of the existence in these nuclei of long isomeric periods.

The observed disturbances in the monotonous increase of δ (h, γ) with A may be accounted for by fluctuations of the value of energy between neutron and nucleus, on account of which the factor $\Gamma_\gamma/\Sigma\Gamma_\alpha$ will be different for two neighbouring nuclei. And if we assume—which we seem to be entitled to do—that the binding energy of the neutron—nucleus varies within 6—8 MeV, this will account in some measure for the cases, where cross-sections for the capture of fast neutrons by heavy nuclei were observed to grow irregularly with atomic weight. With all that a strong forbiddance for the (h, γ) process may possibly exist for certain nuclei. Thus, in bismuth the (h, γ) process should lead to the familiar RaE. All my attempts to produce a detectable transmutation of Bi^{209} (h, γ) Bi^{210} (RaE) by bombardment with (d, d) neutrons, like the many similar attempts made earlier [5], were a failure, though with deutons of 5 MeV the reaction Bi^{209} (d, p) Bi^{210} (RaE) is known to proceed without difficulty [6].

I am indebted to I. V. Kurchatov, Member of the Academy, and I. I. Lukirsky, Corresponding Member of the Academy, for discussion of the results, and to the workers of the cyclotron laboratory for assistance in the bombardment experiments.

Radium Institute.
Academy of Sciences of the USSR.

Received
6. XI. 1944.

REFERENCES

[1] V. Weisskopf and D. Ewing, Phys. Rev., **57**, 472 (1940). [2] V. Weisskopf, Phys. Rev., **52**, 295 (1937). [3] H. von Halban and L. Kowarski, Nature, **142**, 392 (1938). [4] G. Griffiths, Proc. Roy. Soc., A, **170**, 513 (1939). [5] E. Amaldi, O. d'Agostino, E. Fermi, B. Pontecorvo, F. Rasetti and E. Segré, Proc. Roy. Soc., A, **149**, 522 (1935); J. McLennan, L. Grimmet and J. Read, Nature, **135**, 505 (1935); L. Sosnowski, C. R., **200**, 1027 (1935); P. Preiswerk and H. von Halban, C. R., **201**, 722 (1935); E. Andersen, Nature, **137**, 457 (1936). [6] J. Livingood, Phys. Rev., **50**, 425 (1936).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

TECHNICAL PHYSICS

ON THE STRENGTH AND DUCTILITY OF QUENCHED AND TEMPERED STEEL

By J. B. FRIEDMAN and T. A. VOLODINA

(Communicated by A. F. Joffé, Member of the Academy, 20. XI. 1944)

The importance of quenched and tempered steel in modern engineering is tremendous. Until lately the fundamental characteristics of these steels were: ultimate strength σ_B and Brinell hardness H_B, connected with it, as characteristics of strength; relative elongation δ_{10} or δ_5, and reduction of area ψ as characteristics of ductility, and impact toughness a_k as characteristics of toughness. At the same time the now little employed torsion method enables the detection of a number of new properties of steels, that are otherwise undetectable, or do not lend themselves to quantitative measurement during tension tests.

In many cases, especially in quenched and low-tempered steels, the cited ductility and toughness characteristics have such small values (close to zero) that such materials are usually qualified as completely brittle. At the same time the modern theory of deformation shows that such a conclusion is not accurate, for the brittleness of such steels is inherent in them only in definite modes of loading (tension, bending) in which the part of the tensile stresses and strains is sufficiently great(1).

The very possibility of measuring the Brinell hardness (*i. e.* plastic deformation in the form of an indentation on the surface) already shows that, as distinct from still more brittle materials (for example, many glasses and silicates), quenched and low-tempered steels in definite states of stress are capable of quite appreciable plastic deformation (see also(2)). Hence it is necessary to find an exact quantitative method of studying the ductility of such steels.

1. As Davidenkov(3) pointed out for the first time while discussing the results of the well-known experiments of Luerssen and Green(4), the most suitable method for such a study is the method of torsion of a cylindrical bar, which had already been employed by Emmons during a study of instrumental steels(5). During torsion the ratio of the maximum tensile stress to the maximum shearing stress is twice as small as during tension, and indeed, many steels which are almost completely brittle during tension showed appreciable plastic deformation during torsion. In point of illustration deformation curves* for steel of the chromansil type (0.38 per cent C, 1.06 per cent Cr, 0.89 per cent Mn, 1.06 per cent Si) quenched at 880° C in oil without tempering are given in Fig. 1. Whereas during tension (curve P) the steel is completely brittle (elongation is practically equal to zero), during torsion (curve K, Fig. 1) the very same steel shows appreciable ductility (the plastic shear is 30 per cent, which upon recomputation in terms of plastic elongation gives $e=20$ per cent).

* In all experiments, 4 or 5 specimens were tested at one point.

Computation of the stresses and strains during torsion is carried out with the formulae of Ludwik and Nadai (6).

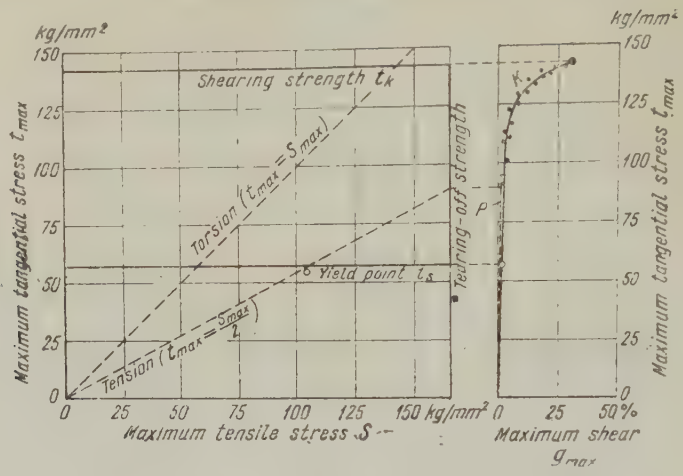


Fig. 1. Diagram of the mechanical state of quenched and tempered chromansil steel

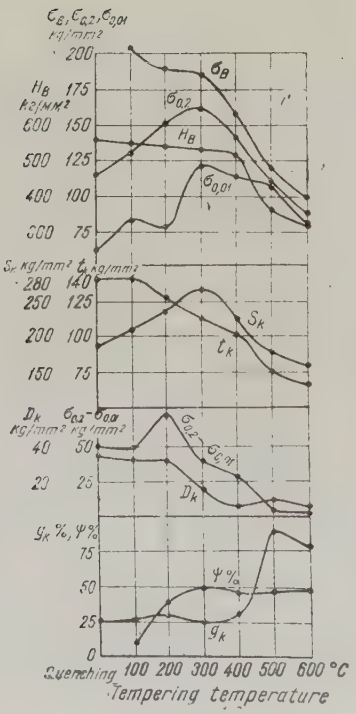


Fig. 2. The mechanical properties of chromansil steel at different tempering temperatures. σ_B, ultimate tensile in kg/mm ; $\sigma_{0.2}$, the proof stress in kg/mm²; $\sigma_{0.01}$, limit of elasticity in kg/mm²; H_1, Brinell hardness in kg/mm²; S_k, true rupture stress in kg/mm²; t_k, shearing strength in kg/m²; $\sigma_{0.2} - \sigma_{0.01}$, difference, approximately characterizing work-hardening in the initial stage of deformation: D_k, coefficient of work-hardening (plasticity modulus) at high degrees of deformation in kg/mm² per 1 per cent deformation; ψ, reduction of area in per cent; g_k, maximum plastic shear during torsion in per cent

Thus, the plasticity of hard quenched steels, which are completely brittle during tension, may be easily and sufficiently reliably evaluated from torsion

experiments, which makes it possible to introduce the quantitative characteristics of the ductility in tests of instrumental, ball-bearing and similar steels*.

2. No less essential are the advantages presented by the torsion method in studying steels that are also highly plastic during tension (for example, after annealing or high tempering), which had already been pointed out by Sauveur(7) and Bain(8). Elongation is known to be unsuitable in principle as an index of the

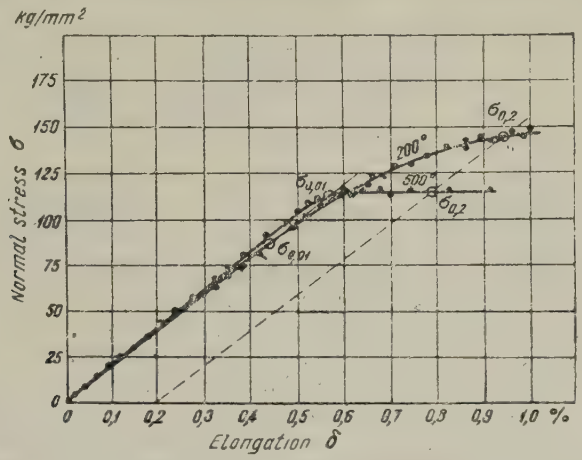


Fig. 3. Deformation curves for chromansil steel after tempering at 200 and 500°. At very small strains (0.01 percent) tempering at 200° gives less strength than tempering at 500°; at large deformations (0.2 per cent) the relation is reverse

ductility(9). However, reduction of area as well gives strongly distorted results, when the average and not the maximum ductility is measured in a non-uniform stressed and strained state.

In Fig. 2 are given various mechanical properties of steel of the above-mentioned composition as a function of the tempering temperature **. Quite an appreciable difference is obtained between the mode of change in the ductility at tension ψ and the maximum shear of torsion g_k. Both for the mentioned steel and a number of other steels $g_p \gg g_k$, whereas thanks to the non-uniformity of the deformed state in the neck of the specimen under tension, it would be natural to expect the opposite. So far no explanation for this phenomenon has been provided. It is possible that the different rotation of the crystallographic planes during tension and torsion has an essential influence here.

3. From Fig. 2 it is seen further that evaluation of the resistance to plastic deformation *** as a function of the magnitude of the strain at which the strength is determined, may lead to exactly opposite results. Thus, in Fig. 3, for example, a comparison is made of deformation curves for chromansil steel after tempering at 200° and 500°. Specimens tempered at 500° are stronger at very small strains ($\sigma_{0.01}$), and those tempered at 200°, at large deformations ($\sigma_{0.2}$).

This can also be clearly seen in Fig. 2. Whereas the σ_B and H_B curves which characterize the resistance of plastic deformation at appreciable strains, gradually fall off with an increase in the tempering temperature (the higher the tempering temperature, the less «strong» the steel), the $\sigma_{0.2}$ and $\sigma_{0.01}$ curves give a maximum at tempering temperatures of the order of 300—400°, while lower tempering tem-

* In the testing of highly brittle steels, during torsion one may observe rupture by tearing off along a line on a 45° incline to the axis of the bar. For these materials transition is necessary to still «milder» modes of loading, for example, to compression.

** For the sake of comparison the reduction of area may be obtained in terms of the true plastic shear g_p with the formula $g_p = 1.5 \ln \frac{1}{1-\psi}$.

*** By strength in the broad sense of the word we have in mind resistance to deformation, plastic as well as causing failure.

peratures give lower «strength», evaluated from $\sigma_{0.2}$ and $\sigma_{0.01}$. That is why in speaking of the strength of a material (in the given case of tempered steels), it is necessary in order to avoid contradictions to indicate the value of the deformation at which this «strength» is compared.

The results given in Figs. 2 and 3 support the views of Kishkin ([10]) regarding the low hardness of martensite, when this hardness is measured at appreciably smaller strains than in the usual Brinell tests.

4. Neither reduction of area ψ (the abscissa of the end-point of the deformation diagram during tension), nor the «true breaking strength» S_k (the ordinate of the end-point on the same diagram), *i. e.* the true tensile strain at the moment of rupture, exactly reflect the maximum resistance shown by the metal to destruction.

It is known that most plastic materials rupture during tension from shearing strains([1]). If there were no neck during tension, one should expect from the ratio of the normal and tangential stress the relation $S_k=2t_k$, where t_k is the shearing strength. Such a relation is actually observed, for example, in pressed magnesium which does not form a neck during tension.

At the same time from Fig. 2 it is seen that $S_k>2t_k$, which may be explained only by the influence of the triaxial state of stress set up by the neck in the internal zones of the specimen([1]). From what has been said it follows that in plastic metals which form a neck during tension, neither reduction of area ψ, nor the true breaking strength S_k may serve as very exact characteristics of the mechanical properties of the material, since they depend upon the highly complex stressed and strained states that arise in the neck of a specimen subjected to tension. Finally, the true deformation diagrams—the flow curves (true curve, investigated heretofore almost exclusively during tension) are distorted([11]) after the formation of a neck for the above-mentioned reasons.

Not only tension, but also torsion holds forth great interest as the basic method of obtaining the characteristics of maximum ductility and strength, and also the true deformation curves for plastic metals. As for the resistance to tearing-off (brittle strength) which is frequently not detected during torsion and at the same time may noticeably affect the properties of the material under tension, this resistance must be evaluated separately([1,6]).

In many cases (in brittle materials) the resistance to tearing-off may be conveniently measured at bending.

All-Union Scientific Research Institute
of Aviation Materials.

Received
20. XI. 1944.

REFERENCES

[1] Я. Б. Фридман, Единая теория прочности материалов, М., 1943. [2] Л. В. Альтшуллер, Н. А. Решеткина, А. Г. Спектор и Б. А. Цукерман, ЖТФ, XIII, No. 6 (1943). [3] N. N. Davidenkov, Metal Progress, **30**, 55 (1936). [4] G. V. Luerssen and O. V. Green, ASTM, **33**, II, 315 (1933); ASM, **23**, 861 (1935). [5] W. H. Emmons, ASTM, **31**, II, 47 (1931). [6] Я. Б. Фридман, Техника возд. флота. No. 1 (1944). [7] A. Sauveur, ASTM, **39** (1939). [8] E. Bain, Functions of Alloying Elements in Steel, pp. 98, 1939. [9] H. Gillet, Metal Progress, Sept. (1939). [10] С. Т. Кишкин, Техн. возд. флота, № 11 (1943). [11] Г. А. Смирнов-Аляев, Упругость и пластичность, сборн. работ Ленингр. ун-та, вып. II, стр. 298, 1935.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

TECHNICAL PHYSICS

ELECTRON MICROSCOPE AND INVESTIGATION OF THE STRUCTURE OF CERAMIC MATERIALS

By I. I. KITAIGORODSKY

(Communicated by A. F. Joffé, Member of the Academy, 11. XI. 1944)

Electron rays are absorbed by matter very readily, and hence, the very thinnest layers of matter are opaque to them. That is why the majority of photographs produced with the electron microscope are shadow pictures.

However, one may notice on them a gradual thinning out of the particles of the object towards its edge—a transition from shadow to penumbra and thence to light.

Photographs of this type give a reliable information as to the shape and dimensions of the particles of the substance.

In an investigation of ceramic samples which we had undertaken it was important to establish whether the particles of the powder are intimately enveloped by a vitreous substance, or whether glass gathers into separate little particles, just as the crystalline mass gathers into crystallites. This problem can be solved by studying the shadow plotographs made by the electron microscope. Indeed, the shadow picture of a crystal grain, even if it has been crushed, should always be bounded by straight lines. On the contrary, the boundary of the vitreous phase should be of irregular shape.

By means of the electron microscope we photorgaphed a number of powders without glass and analogous powders with glass. The powders investigated were: a) electrocorundum, b) electrocorundum with glass, c) tungsten, d) tungsten with glass, e) tungsten carbide, and f) tungsten carbide with glass.

The resulting photographs showed that:

1) In the majority of cases shadow plotorgaphs afforded quite a positive means of distinguishing the vitreous phase from crystalline and

2) in the case we studied the glass phase and the crystalline grains were intimately bound together, *viz.* a vast majority of crystalline grains were surrounded by a film of the glass phase.

In order to be able to draw conclusions of any kind as to the influence of the glass phase upon the structure of the ceramic material, it is necessary to study under the electron microscope not one, but several systems, since such cases may be met with, where lack of experience in studying that particular kind of electronic photographs will prevent from drawing any definite conclusions as to the structure of the material.

As examples we give the photographs of pure tungsten carbide (Fig. 1a) and of tungsten carbide with glass (Fig. 1b). The distinguishing feature of the pure carbide is the monotony of particles, evidently, accumulations of tiny crystallites; great numbers of them can be observed in all the snaps. This specific pecu-

liarity disappears in the photographs of the system tungsten carbide—glass: there was not one picture of this sample which had on it the tiny particles found in the case of the pure carbide. A comparison of these photographs gives ample grounds to conclude that the tiny tungsten carbide particles are cemented with the glass into large aggregates.

A comparison of the photographs of tungsten with glass (Fig. 2a) and pure tungsten (Fig. 2b) shows that the basic grain is enveloped in a thin layer of glass. In Fig. 2b all the grains are distinctly defined, and the glass phase is altogether absent.

The investigations with the electron microscope were carried out in the Physical Laboratory of the All-Union Institute of Experimental Medicine by B. M. Isaev and A. I. Kitaigorodsky, to whom the author expresses due gratitude.

Received
11. XI. 1944.

Fig. 1a

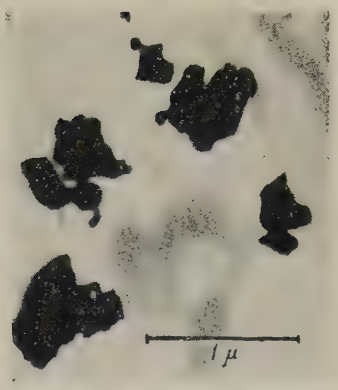


Fig. 1b

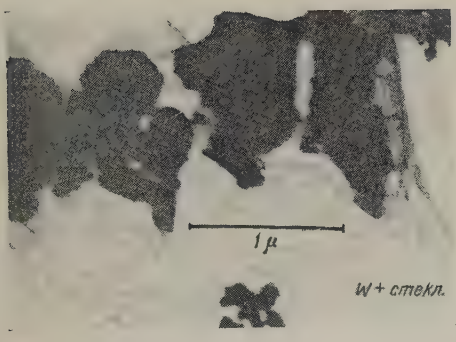


Fig. 2a

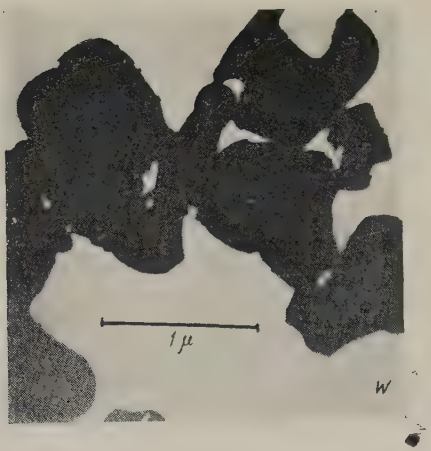
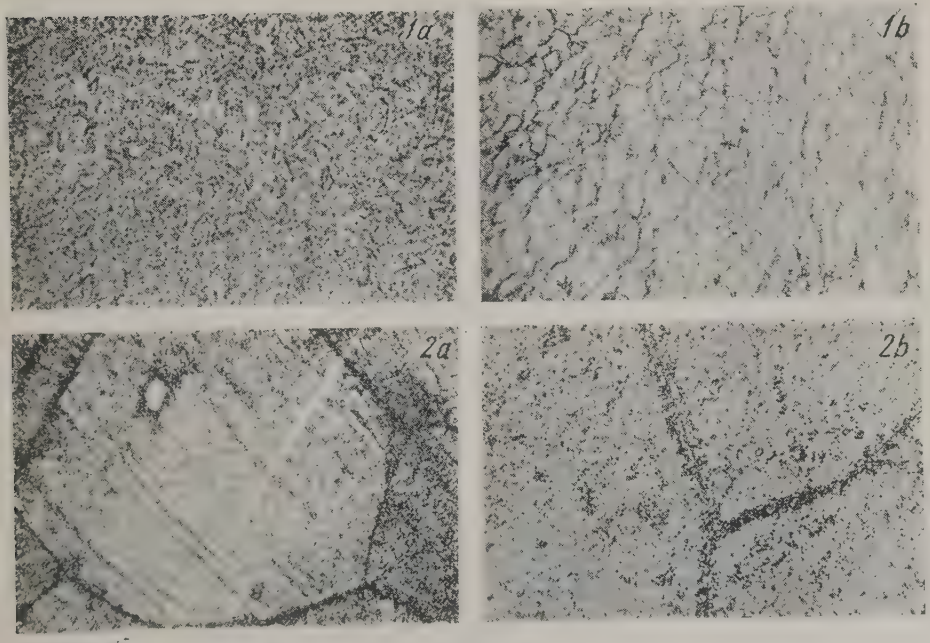


Fig. 2b

G. V. Akimov and *Z. L. Mintz*



Microstructure of *EI*-69 steel tempered at 1050° (1a, 1b) and at 1250° (2a, 2b) after aging at 800° for a duration of 30 mins. (1a, 2a) and 50 hrs. (1b, 2b), ×500

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

PHYSICAL CHEMISTRY

STRUCTURE OF AUSTENITIC HEAT-RESISTANT STEEL

PROCESSES INVOLVED IN THE AGING OF HARDENED EI-69 STEEL

By G. V. AKIMOV, Corresponding Member of the Academy, and Z. L. MINTZ

The EI-69 steel samples used in this study of the aging of hardened austenitic steels were the same as those used in the first part of this work (1).

A comparison was made of the structures formed during aging for two tempering temperatures—one of them, the usual one, 1050° (fine grain); the other, somewhat higher, 1250° (coarse grain). Aging was carried out at temperatures of 700—850°, the exposure intervals lasting from 30 mins. to 50 hrs.

The results of this investigation are given in the figure for an aging temperature of 800° and the extreme intervals of exposure—30 mins. and 50 hrs. The results obtained with the other temperatures and for intermediate durations were analogous to these.

As can be seen from the accompanying microstructures, aging processes in steel that has been tempered at 1050° differ greatly from those in steel tempered at 1250°. During the aging of fine-grained steel (tempered at 1050°) the carbides are deposited evenly throughout the whole grain. These carbides are rounded in shape and of approximately the same dimensions as the carbides which remained undissolved in the steel when the latter was heated before tempering. This is due to the fact that before being aged the structure contains some remnants of undissolved carbides, which readily become crystallization centres for fresh carbides. These new, fine-grained carbides, evidently, coagulate easily, rapidly forming carbides of microscopic size. While the carbides are being deposited, the boundaries of the grains are not marked in any way.

In steel tempered at 1250° the process of aging proceeds differently. First of all, the carbides precipitate in a very minutely dispersed form. At 750° and with short exposure (30 mins.) the only manifestation of the precipitation is a general change in the etchibility of the ground and a blackening of the boundaries of the grains, twins and slide lines. Even at an aging temperature of 800° with the same duration individual carbides are still difficult to distinguish. With a duration of 50 hrs. at 800° the coagulated carbides can be distinguished as parallel chains only along the grain boundaries, where the density of precipitated carbides is, apparently, very high. The grain itself, however, is filled with very minute carbide deposits. Thus, the carbides in steel tempered at 1250° become thermally stable and coagulate slowly, chiefly in those places where the density of carbide deposits is high.

This difference in the precipitation of carbides may be explained as follows: in EI-69 steel, carbides of various types (chrome-iron, chrome-iron-tungsten, and tungsten) may form. The thermal stabilities of these carbides differ. Naturally, the less stable carbides will be first to pass into solid solution when heated. Thus, if the carbides are not dissolved completely when the steel is tempered, those which will later precipitate out of the solid solution in the course of aging are thermally the less stable carbides. If, on the other hand, all the carbides, in-

cluding those with the greatest thermal stability, pass into the solid solution, the carbides which precipitate out during aging will be of different composition, and will be much more stable thermally.

The difference in the aging of steels tempered at higher and lower temperatures explains the difference in mechanical properties observed at high temperatures.

From this standpoint, the high thermal stability and low plasticity at high temperatures of coarse-grained steels is due not so much to the presence of coarse grains as to the formation of extremely fine-grained carbides of high thermal stability; the preferential precipitation of carbides along the boundaries of the austenite grains during the first stages of aging probably plays a considerable rôle in the hardening of the austenite grains

Laboratory of Physics of Metals.
All-Union Institute of Aircraft Materials.

Received
27. I. 1945.

REFERENCES

[1] Z. L. Mintz and G. W. Akimov, C. R. Acad. Sci. URSS, XLVIII, No. 3 (1945).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

PHYSICAL CHEMISTRY

ON THE DEPENDENCE OF THE BURNING VELOCITY OF SMOKELESS POWDERS UPON THE PRESSURE

By K. K. ANDREEV and A. P. BAKEEV

(Communicated by N. N. Semenov, Member of the Academy, 5. XI. 1944)

It is known that the burning velocity of smokeless powders increases with the pressure. The form of this dependence, however, is still uncertain. In the general case the dependence of the burning velocity upon the pressure may be expressed by the formula

$$u = A + Bp^{\nu} \tag{1}$$

However, in internal ballistics of late quite extensive use ([1]) has been made of the more simple formula

$$u = Bp \tag{2}$$

Thus, its supporters assume that A is equal to zero, and ν is equal to unity.

On the other hand, in researches on the burning of powders in rockets it has been shown that formula (2) leads to disagreement with the facts([2]). From a comparison of the expressions for the rate of flow of gases through a nozzle and for the amount of gas formed as a result of burning it follows that if the burning velocity is directly proportional to the pressure, it is possible for the amount of gas formed and let out to be equal for a given nozzle at any pressure within the chamber. At the same time it is known that for powder rockets there is a definite stable value of the pressure within the chamber for each given charge and nozzle when the regime is stationary. This may be the case if $\nu \geqslant 1$, or if $\nu = 1$, but $A \neq 0$.

It should be emphasized that none of the experimental data are based on direct determination of the value of the burning velocity at constant pressure, but all have been obtained by deduction from various curves characterizing the course of burning at variable pressure, increasing from 1 to 1000 atm. and higher. Apart from the considerable complexity of the burning process under these conditions, the derivation of a general formula for the dependence of the burning velocity upon the pressure would be legitimate only if this dependence remained invariable throughout the whole pressure interval. This, however, has never been proved.

What is more, theoretical considerations give us grounds for doubting the invariability of the mentioned dependence in a broad range of pressures, inasmuch as the very mechanism of the burning process and its characteristics may change essentially with a change in pressure. Thus, at low pressures, when the heated layer of powder is thick, a considerable rôle may be played by a reaction, in the condensed phase, which is influenced by the pressure, in the main, only through the magnitude of the thickness of the heated layer. Inasmuch as this thickness decreases with a rise in pressure, the absolute value of the reaction velocity in the condensed phase (the amount of powder that decomposes in unit time) will accordingly decrease. The total burning velocity will thus be the sum of two functions—decreasing (reaction in the condensed phase) and increasing (reaction

in the gas phase). Furthermore, with an increase in the pressure at which combustion is taking place, the over-all equation of the reaction changes in an ever greater and greater degree. At low pressures the bulk of the nitrogen is evolved in the form of nitrogen oxide; at high pressures, it is practically all evolved in the elementary form. The quantities of hydrogen, water, carbon monoxide and methane change accordingly, approaching the equilibrium values at the temperature and pressure attained. The indicated change in the over-all combustion equation signifies also an appreciable change in the heat effect, and hence in the temperature of combustion as well, which in turn should affect the combustion velocity.

Computation of the influence of all these factors upon the value of the burning velocity is quite complicated, and therefore the reliability of the theoretical calculations of the velocity and its dependence upon the pressure is very limited, at least for the range of pressure within which the influence of the mentioned factors cannot be disregarded. Only direct experiment—a determination of the burning velocity at constant pressure—can furnish the most authentic data in this direction.

Researches of such a nature were launched by K. K. Andreev in collaboration with A. E. Varga in 1937 ([3]). A study was made of the burning velocity at constant pressure of nitroglycerol powder with 28 per cent nitroglycerine in the pressure interval from the lowest, at which combustion can take place, to 1600 mm Hg. The experiments were carried out with powder cylinders wrapped in asbestos at elevated temperatures (85 and 115°C), since at ordinary temperatures the powder does not burn under the given conditions. The data obtained showed that in the pressure range from ~800 to 1600 mm Hg the burning velocity increases linearly with the pressure in keeping with the formula $u = A + Bp$.

The data, however, were for a very low pressure, and the pressure range investigated was relatively very narrow (~1 atm.). This made it uncertain whether the dependence obtained could be extended to higher pressures without changing the numerical values of the coefficients A and B.

From these considerations the investigations were expanded to embrace higher pressures. In the present paper we will examine the results obtained at pressures of up to 50 kg/cm^2. The experiments were carried out in a thick-walled steel bomb having the following specifications: internal diameter, 60 mm; length, 460 mm; volume, 1.3 litres. The bomb was connected by two copper tubes, with an internal diameter of 16 mm, to two steel tanks, each having a capacity of 150 litres. Before the experiment the tanks and bomb were filled with air at the same pressure at which combustion was to take place. Thanks to the large volume of the system and the total absence or insignificance of leaks, burning proceeded practically at constant pressure. When there were slight fluctuations, the mean value was taken from the pressure at the beginning and the end of combustion. The powder was employed in the form of solid cylinders, 5 mm in diameter and 100 mm long. Then an asbestos rope was wrapped around the cylinder to a thickness of ~2 mm, and the whole fitted more or less snugly into a glass tube. Burning was followed through a thick organic glass window fitted into the wall of the bomb. The end of combustion was indicated by the flash of a small quantity of black powder placed under the lower end of the cylinder.

Experiments were carried out with four kinds of powder:

I. Nitroglycerol powder of a composition previously investigated by Andreev and Varga.

II and III. Nitroglycerol powders close in composition to I.

IV. Nitrocotton powder.

The experimental results may be represented by the following formulae:

$$u = 3.6 + 1.202\ p^{0.84} \quad \text{(powder I)}$$
$$u = 3.1 + 1.202\ p^{0.90} \quad \text{(powder II)}$$
$$u = 1.4 + 0.85\ p \quad \text{(powder III)}$$
$$u = 2.2 + 0.447\ p^{1.12} \quad \text{(powder IV)}$$

Graphically they are described by the curves in Figs. 1 (powders I and II) and 2 (powders III and IV).

In agreement with the conclusions of Andreev and Varga, a certain pressure limit exists ($\sim$2 kg/cm^2) below which combustion does not take place in the given experimental conditions. At large pressures the burning velocity increases with pressure.

For all four powders investigated the term A, which does not depend upon the pressure, should enter the expression for the dependence of the burning velocity

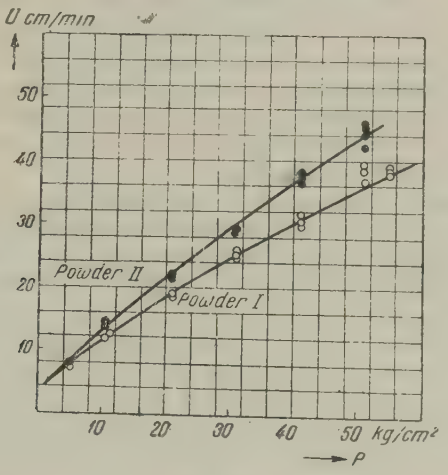


Fig. 1

upon the pressure. At $p = 5$ kg/cm^2 this term is 1/4 to 1/2 of the burning velocity.

The change in velocity with pressure is nearly linear in the investigated interval. The graphs give the values of ν from 0.84 for powder I to 1.12 for nitrocotton powder IV.

If the relations obtained are compared with the results of a determination of the characteristics of the powders in a manometric gauge at a charge density

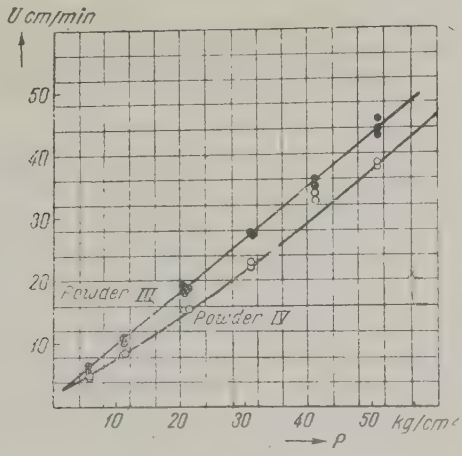


Fig. 2

$\Delta = 0.20$ ($P_{max} = 1500$—2000 kg/cm^2), it may be concluded that at high pressures the exponent for powder I is somewhat larger than 0.84, and for powders IV, III and II, less than 1.12, 1 and 0.9, respectively. Evidently, the mean value of this exponent for diverse powders in a broad range of pressures lies between 0.84 and 0.90. It may hence be asserted that ν is very close to unity, and in any case incom-

parably closer to unity than to 1/2, the value proposed of late in several works on the basis of theoretical considerations [4].

Thus, smokeless powders at pressures of 2—50 kg/cm² give approximately the same law of burning $u = A + Bp$, as was previously established by Andreev and co-workers for nitroglycerol powders and a number of secondary explosives at pressures up to 2 kg/cm².

If we proceed from the premises Zeldovich and Belayev worked from in deriving the formula for the combustion velocity, this means that the reaction determining the burning velocity is bimolecular. In this case comparison of the characteristics of decomposition during combustion with the characteristics of slow monomolecular thermal decomposition [4] hardly has any actual meaning. This also emphasizes the arbitrary and limited meaning of the simple, essentially empirical formulae of the type $u = A + Bp$ with the constant ν, approximately applicable only to those relatively narrow pressure intervals for which the coefficients A, B and ν have been experimentally established.

In applying the results obtained to the burning of a charge in a fire-arm, it should be kept in mind that in these conditions the found relation may, even within the limits of the pressure interval for which they were established, be greatly distorted by diverse factors. Thus, for example, in the case of tubular powder of a great length and small diameter of the bore, the velocity of flow of the gases in the bore undoubtedly has an effect, although neither it nor the increase of pressure in the bore have as yet been directly established by experiment.

Further, at a certain depth, determined by the thickness of the heated layer, combustion of the powder at the beginning and end may not be stationary. In exactly the same way the burning velocity of the powder tube at the outside surface, *i. e.* with a decreasing area of the front, may differ from the burning velocity at the butt end of the cylinder, *i. e.* when the area of the front is constant. In the case of tubular powder, it is true, the influence of the change in area of the combustion front should be mutually more or less compensated, especially at raised pressures, when this influence in general is less appreciable.

The D. I. Mendeleyev Institute
of Chemical Technology.

Received
5. XI. 1944.

REFERENCES

[1] М. Е. Серебряков, К. К. Гретен, Г. В. Оппоков, Внутренняя баллистика, 1939, стр. 109. [2] М. Руа, О полезном действии и условиях применения ракетных аппаратов, 1936, стр. 28. [3] М. Патри, Горение и детонация взрывчатых веществ, 1938, стр. 161. [4] А. Ф. Беляев, ЖФХ, 14, 1009 (1940). [5] К. К. Андреев и сотр., Сборн. статей по теории взрывчатых веществ, 1940, стр. 39.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS]
1945. Volume XLVIII, № 8

BIOCHEMISTRY

ACTIVATION OF CATALASE BY BIOGENIC STIMULANTS

By A. V. BLAGOVESHCHENSKY and A. J. KOLOGRIVOVA

(Communicated by A. A. Richter, Member of the Academy, 23. II. 1945)

According to Filatov, when the cells are subjected to conditions approaching lethal ones, for example, are chilled during a long time at a temperature slightly above 0°, certain substances which counteract destruction are formed by them. These substances were termed by Filatov factors of resistance, or biogenic stimulants.

The first attempt to give these phenomena of stimulation biochemical interpretation was made by one of us as early as in 1939. An experimental study on the activeness and quality of catalase in cucumber seedlings chilled at $+1°$ during a week has shown that the activeness of catalase was increased as compared to the control, the thermal coefficients being at the same time reduced, *i. e.* an improvement of the quality of the catalase, or a rise of its energy level had taken place.

The constants of the rate of reaction of splitting of hydrogen peroxide at 15° increased from $29.7 \cdot 10^{-4}$ in the control to $43.2 \cdot 10^{-4}$ in the chilled, and from $54.3 \cdot 10^{-4}$ at 25° in the control to $63.0 \cdot 10^{-4}$ in the chilled. The value of Vant Hoff's coefficient has consequently dropped from 1.83 in the control to 1.46 in the chilled, and the value of Arrhenius' coefficient has dropped from 10300 gcal to 6400 gcal, respectively.

The results of these experiments gave, however, no information with regard to the question, whether chilling results in an alteration of the protein of the catalase itself, or the primary change consists in the formation of some special substances which act upon the enzyme and increase its activeness and improve its quality. A positive answer to this question in the latter sense was obtained from the experiments on the extracts from the seedlings of *Phaseolus aureus* Roxb., which were chilled at $+2°$ during 10 days.

The seedlings were ground, the juice filtered, and the proteins coagulated by boiling. The liquid was sterilized in an autoclave at 120° and contained neither proteins nor enzymes; in particular, it was absolutely unable to decompose hydrogen peroxyde. The therapeutical effect of this extract was tried by V. V. Skorodinskaya and proved to produce well-expressed beneficial results (in treating the «Pendina ulcer», a skin leishmaniosis).

The resting seeds and seedlings of *Phaseolus aureus* Roxb. of various age were ground down in a buffer mixture with pH = 7.0. The resulting suspension was divided into two portions, one of which was added the preparation of the stimulant, and the other, water. The activeness of the catalase was studied gasometrically. The constants of the rate of the reaction were computed from the averages obtained from three parallel determinations. The computation was done

after the formula: $k = \frac{1}{t} \ln \frac{a}{a-x}$. From the results obtained the other values were computed:

$$Q_{10} = \frac{k_2}{k_1} \quad \text{and} \quad \mu = R \ln Q_{10} \frac{T_1 T_2}{10} \text{ gcal.}$$

Experiment 1. Weighed portions of fine-ground germinated seeds (250 mg) were mixed with 25 cm³ of the buffer mixture. In the control series 2 cm³ of water was added to this mixture, while in the experimental one there was added 2 cm³ of the stimulant. For each determination 5 cm³ was taken. The amount of oxygen liberated after hydrogen peroxide is completely decomposed: $a = 35.5$ cm³. Readings were taken every 30 seconds. The temperature was 15 and 25° (Table 1).

Table 1

Kinetics of the Action of Catalase of Seeds under the Influence of Biogenic Stimulants

t''	Liberation of O_2 in cm³				$k \cdot 10^5$			
	Control		Experiment		Control		Experiment	
	15°	25°	15°	25°	15°	25°	15°	25°
30	7.2	11.4	8.3	11.6	754	1289	886	1317
60	11.6	18.3	12.4	18.0	658	1168	735	1177
90	15.0	21.3	16.2	21.2	610	1017	676	1009
120	17.8	23.2	19.1	22.8	579	882	643	855
150	20.0	23.9	21.5	23.8	552	745	620	739
			Average		631	1020	708	1019
			Q_{16}		1.62		1.44	
			μ gcal.		8400		6400	

Experiment 2. 3 g of five-day-old seedlings which were finely ground with 25 cm³ of buffer solution, 5 cm³ of the stimulant (or of water). $a = 30.55$ cm³ (Table 2).

Table 2

t''	Liberation of O_2 in cm³						$k \cdot 10^5$					
	Control			Experiment			Control			Experiment		
	10°	15°	20°	10°	15°	20°	10°	15°	20°	10°	15°	20°
30	8.85	9.3	10.35	10.25	10.05	11.55	1136	1283	1377	1358	1327	1572
60	12.8	14.25	16.0	14.35	15.15	16.70	923	1043	1232	1159	1137	1313
90	15.8	17.25	18.75	16.30	17.90	19.15	806	920	1056	864	976	1094
120	17.65	19.05	20.35	17.85	19.30	20.40	716	811	910	729	829	914
150	18.7	19.85	21.15	18.90	20.40	20.95	629	696	783	640	732	768
180	19.31	20.3	21.6	19.35	20.90	21.45	—	—	—	—	—	—
				Average			842	951	1072	950	1000	1132
				Q_{10}			1.27			1.19		
				μ gcal.			3900			2900		

Below we shall present only the average values of $k \cdot 10^5$ of several experiments (Table 3). The data relate to resting seeds (experiment 3) and to seeds which were soaked during one night (experiments 4 and 5).

Table 3

Constants of Rate and Thermal Coefficients of Catalase under the Influence of Biogenic Stimulants

No. of experiment	$k \cdot 10^5$				Q_{10}		μ gcal	
	Control		Experiment		Control	Experiment	Control	Experiment
	15°	25°	15°	25°	15°	25°	15°	25°
3	438.2	773.3	502.2	770.5	1.76	1.53	9600	7200
4	280.8	534.2	305.0	563.9	1.90	1.85	10900	10500
5	288.5	573.9	301.7	410.3	1.99	1.36	11700	5200

In all the above experiments, as well as in the experiment with catalase, the activeness of the enzyme had thus grown under the influence of the stimulant; at the same time its quality was also improved. One could suggest that the improvement of the quality was only an illusory one, since at 25° there was observed in the majority of experiments a certain reduction of the constant of rate of reaction in the presence of the stimulant, which could be interpreted as the breakdown

Table 4

Thermal Inactivation of Catalase

t''	Without the stimulant		With the stimulant	
	15°	25°	15°	25°
	Beginning of experiment			
30	331.1	890.9	341.2	514.5
60	293.8	573.8	299.9	446.3
90	278.6	522.9	352.6	366.1
120	268.5	459.0	259.0	379.3
150	255.3	423.1	255.9	345.2
Average	285.5	573.9	301.7	410.3
	30 minutes stay			
30	308.2	401.1	321.2	311.4
60	271.6	354.5	310.8	292.8
90	250.4	348.5	281.4	270.4
120	250.6	316.8	272.0	263.8
150	232.5	296.7	255.8	245.3
Average	262.7	343.5	288.2	276.7

(Continued)

t°	Without the stimulant		With the stimulant	
	15°	25°	15°	25°
		1 hour stay		
30	207.8	374.1	248.3	342.7
60	197.7	397.7	241.4	344.2
90	190.8	371.6	213.9	327.9
120	174.5	338.4	201.7	317.1
150	164.1	317.4	191.7	284.5
Average	187.0	359.8	219.4	323.3
		2 hours stay		
30	207.8	505.2	227.7	332.0
60	203.1	479.0	241.4	350.3
90	194.7	459.8	225.6	336.6
120	183.3	418.3	217.3	313.6
150	168.9	396.0	204.5	299.3
Average	191.6	451.7	223.3	326.4

of the enzyme at this temperature. This circumstance made us undertake a special investigation of the thermostability of catalase in the absence and in the presence of biogenic stimulants.

Seeds of *Phaseolus aureus* which were swollen in water were carefully ground with the buffer solution and mixed with the solution of the stimulant (in the control with an equal quantity of water, instend). The determinations of the activeness were made immediately after preparing the mixture and then every 30 minutes, 1 hour and 2 hours, during which the mixture was kept at 15 and 25°.

In table 4 are given the values obtained of the constants of the velocity $k \cdot 10^5$.

From a comparison of the values of the constants of rate for the catalase without the stimulant and in the presence of stimulants we find that in the latter case the enzyme becomes more stable at a temperature of 15°, but its stability is less at 25°, which leaves open the question, whether the quality of the catalase had improved in our experiments.

Received
23. II. 1945.

REFERENCES

[1] В. П. Филатов, Тканевая терапия, Ташкент, 1943.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

BIOCHEMISTRY

INTENSITY OF FIXATION OF ATMOSPHERIC NITROGEN BY *AZOTOBACTER* ON DIFFERENT SOURCES OF CARBON

By M. V. FEDOROV

(Communicated by A. A. Richter, Member of the Academy, 26.XI.1944)

It has been clearly demonstrated in numerous works carried out by different authors that nitrogen fixing bacteria are able to fix considerable quantities of atmospheric nitrogen. In spite of the large number of hypotheses advanced ([1–3]) by various authors, the precise mode of fixing nitrogen in the cultures of the microorganism investigated is not known. The author believes that the material discussed below permits of a new approach to the analysis of this process.

In his experiments the author tried to find out the relation existing between the structure of the organic matter used by nitrogen fixing bacteria as a source of carbon and energy, and the rate of their fixing the nitrogen of the atmosphere. With this aim in view a series of experiments were undertaken in Erlenmeyer flasks with a capacity of 300 cm^3. Each flask was made to contain 100 cm^3 of mineral medium for the nitrogen fixing bacteria, and a quantity of organic matter sufficient to bring the concentration of the latter to $^1/_6$ mole (organic acids were made use of in the form of sodium salts). In all experimental sets a reaction was maintained corresponding to a pH=6.98; this was achieved by adding small portions of Zerensen's phosphate buffer solution, with KH_2PO_4 and Na_2HPO_4 at a varying proportion. Inoculation with *Azotobacter agile* was brought about by adding 1 cm^3 of the bacterial suspension (a bacterial film taken from one test-tube with slanting agar was shaken with 100 cm^3 of sterile tap water). The experiment was carried on during 2 months. Within this period the organic matter was almost everywhere used up by the bacteria. After the completion of the experiment, the total nitrogen was determined after Kjeldahl. The results obtained are summarized in the table.

As may be seen from the figures of the table, there is a definite relation between the structure of the organic matter used up by the bacteria, and the rate of fixation of nitrogen by *Azotobacter agile*. Carbohydrates and compounds closely related to them (glycerine, *etc.*), which are readily oxidized in cultures of *Azotobacter agile*, display a similar productivity both per one gram molecule of the substance and per one gram of the hydrogen contained in this substance. The higher the content of hydrogen, the higher is the rate of nitrogen fixation; thus, with glucose it is 1476, with mannite 1643 mg per 1 gram molecule. The introduction into a molecule of the substance of additional oxygen atoms has an opposite effect, leading to a reduction of the rate of nitrogen fixation, which may easily be seen when comparing the fixation of nitrogen by bacteria grown on succinic, malic and tartaric acids, as well as on lactic and pyroracemic acids. There are certain exceptions to this rule, however. Thus, for example, fixation of nitrogen was found to proceed at a higher rate on glycolic than on acetic acid, although the

Rate of Fixation of Atmospheric Nitrogen by *Azotobacter agile* on
(organic acids used in

Source of carbon	Amount of matter per 100 cm³ of medium, in g	Amount of matter consumed by bacteria during experiment, in g	Atmospheric nitrogen fixed in mg		
			per total substance of an individual culture	per total substance, average	per 1 gram-molecule of substance
Glucose	3.00 3.00	3.00 3.00	25.0 24.22	24.61	1476.6
Mannite	3.03 3.03	3.03 3.03	26.42 28 35	27.39	1643.4
Glycerine	1.53 1.53	1.53 1.53	17.08 16.10	16.59	995.4
Ethyl alcohol	0.767 0.767	0.767 0.767	6.63 4.27	5.45	327.0
Propyl alcohol	1.00 1.00	1.00 1.00	3.90 4.62	4.26	255.6
Acetic acid	1.00 1.00	1.00 1.00	4.06 4.20	4.13	247.8
Glycolic acid	1.27 1.27	1.27 1.27	5.72 5.12	5.12	325.8
Lactic acid	1.50 1.50	1.50 1.50	9.45 9.00	9.23	553.8
Pyroracemic acid	1.47 1.47	1.47 1.47	5.81 6.46	6.14	368.4
Succinic acid	1.97 1.97	1.97 1.97	13.16 15.60	14.38	862.8
Malic acid	2.23 2.23	2.23 2.23	7.98 7.97	7.97	478.2
Fumaric acid	1.93 1.93	0.71 0.53	8.09 8.19	8.14	488.4
Tartaric acid	2.50 2.50	2.50 2.50	8.52 7.96	8.24	494.4
Butyric acid	1.47 1.47	1.47 1.47	16.04 14.44	15.24	914.4
Benzoic acid	2.42 2.42	2.42 2.42	23.24 24.66	23.95	1437.0
Quinic acid	3.25 3.25	3.25 3.25	27.25 29.45	28.35	1701.0

Different Sources of Carbon at a Concentration of 1/6 mole the form of sodium salts)

Chemical structure	Atmospheric nitrogen fixed in mg					Heat of combustion of matter in kcal. per 1 gram-molecule
	per 1 g of H contained in the substance	per 1 g of radical H (without H of hydroxyl and carboxyl)	per 1 g of total C in the substance	per 1 g of radical C (without carboxyl group)	per 1 kcal. of chemical energy of substance used	
$CH_2OH(CHOH)_4CHO$	123.0	211.0	20.5	20.5	2.1	674.0
$CH_2OH(CHOH)_4CH_2OH$	117.4	205.4	22.8	22.8	2.2	728.1
$CH_2OHCHOHCH_2OH$	124.4	198.0	27.6	27.6	2.05	397.4
CH_3CH_2OH	54.7	60.0	13.6	13.6	1.0	328.0
$CH_3CH_2CH_2OH$	32.0	36.0	7.1	7.1	0.53	484.0
CH_3COOH	61.7	83.0	10.3	20.6	1.18	209.6
$CH_2OHCOOH$	81.4	162.0	13.6	27.0	2.0	166.8
$CH_3CHOHCOOH$	92.3	141.0	15.4	23.0	1.70	325.9
$CH_3COCOOH$	92.1	123.0	10.2	15.3	1.40	260.0
CH_2COOH | CH_2COOH	143.8	216.0	18.0	36.0	2.40	356.9
CH_2COOH | $CHOHCOOH$	79.7	160.0	9.97	20.0	1.50	320.1
$CHCOOH$ || $CHCOOH$	122.1	244.0	10.17	20.4	1.53	320.3
$CHOHCOOH$ | $CHOHCOOH$	82.4	247.2	10.3	20.6	1.8	281.1
$CH_3CH_2CH_2COOH$	114.3	130.0	19.0	25.4	1.75	524.8
C_6H_5COOH	239.6	287.2	17.1	20.0	2.0	772.1
$C_6H_7(OH)_4COOH$	141.8	243.0	20.2	23.6	2.0	584.0

reverse should be anticipated. From the above said it should be inferred that the course of oxidation process is probably affected not only by the saturation of a given substance with hydrogen, but abso by the readiness of its subsequent transformation under the influence of oxidizing enzymes, and by many other properties inherent in this substance.

Rather peculiar is the behaviour of substances containing only the primary alcohol group (ethyl and propyl alcohols). Notwithstanding their being high in hydrogen and the presence of a considerable store of chemical energy, they are but poor sources of energy for the nitrogen fixing activity of *Azotobacter*.

This result is possibly correlated with the surface activity of these compounds, since a mere saturation with hydrogen of a given substance does not prevent the fixing of atmospheric nitrogen by the microorganism using it as medium. This is easy to see when comparing the fixing of nitrogen on propyl alcohol and butyric acid.

Lastly, quite apart stand the cyclic compounds. With benzoic acid the productivity of nitrogen fixation is at least twice that observed with glucose, mannite, or organic acids with an open chain. Still higher it is with quinic acid. In so far as the transformation of the latter is connected with the formation of benzene derivatives, it seems that the rate of nitrogen fixation is increased not only by dehydration, but also by the dehydratation of matter, if it is connected with the formation of unsaturated bonds.

The calculation of the atmospheric nitrogen fixed per 1 g of total hydrogen contained in the substance and 1 g of the hydrogen of the radical (without the hydrogen of the carboxyl and hydroxyl group) yields rather inconstant figures, pointing to the absence of any direct relation in this case. Rather incongruent results are also obtained when calculating the amount of the atmospheric nitrogen fixed per 1 g of total carbon contained in the substance; they point to a lack of any chemical regularity in this case, too. More satisfactory are the results obtained from the recalculation per 1 g of the radical's carbon (without carboxyl). In most cases from 20 to 25 mg of atmospheric nitrogen were found to be fixed per 1 g of radical's carbon. A lower rate of production was observed for surface active substances only. The similarity of the rates of nitrogen fixation with different carbon sources should be taken to mean that this process is closely related to the transformations of the carbon of the radical. The calculations of the productivity of nitrogen fixation per 1 cal. of chemical energy of the substance used are found to corroborate this statement. As may be seen from the table, about 2 mg of atmospheric nitrogen are fixed per 1 kcal. in the majority of cases.

The K. A. Timiriazev Agricultural Academy, Moscow.

Received 26. XI. 1944.

REFERENCES

[1] J. Blom, Bioch. Z., 134, 385 (1928). [2] J. Blom, Ztrbl. f. Bakteriol., II Abt., 84, 60 (1931). [3] D. Burk, Bioch. J. (USSR), II, 312 (1937). [4] С. Костычев, А. Рыскальчук и О. Шведова, Тр. отд. с.-х. микроб. Гос. ин-та опыт. агрономии, 1, 91 (1926). [5] S Kostychev and A. Sheloumova, Bull. Acad. Sci. URSS, sér. biol., 662 (1931). [6] O. Loew and Aso. Ztrbl. f. Bakt., II Abt., 22, 452 (1909). [7] J. Stoklasa, ibid., II Abt., 21, 484 (1908). [8] S. Winogradsky, Annal. de l'Inst. Pasteur, 48, 269 (1932).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

BIOCHEMISTRY

DIFFUSION AND MOLECULAR WEIGHT OF LYSOZYME

By A. PASSYNSKY and V. PLASKEYEV

(Communicated by A. N. Bach, Member of the Academy, 8.XII.1944)

As well known, lysozyme is an antibacterial substance which is widespread in nature; thus, for example, it has been discovered in the white of chick egg, in tears, in the secretum of the nasal mucous membrane, in saliva, blood plasma, extracts prepared with certain organs (heart, liver, marrow, kidneys, *etc.*), in milk, turnip, cabbage, horse-radish. Owing to the relatively high concentration of lysozyme in albumin and the convenience of its extraction, egg lysozyme is the basic form of this preparation.

Lysozyme causes the lysis of saprophytic and many other pathogenic microbes; particularly susceptible to its lysogenic effect is *Micrococcus lysodeicticus*. The effect of lysozyme is much similar to that of bacteriophage from which it differs in its rapidity, caused by the absence of latent period, in a lower specificity and a wider range of action. Lysozyme affects not only young but also old bacteria, as well as dead bacterial cells, acting besides within a wider temperature range.

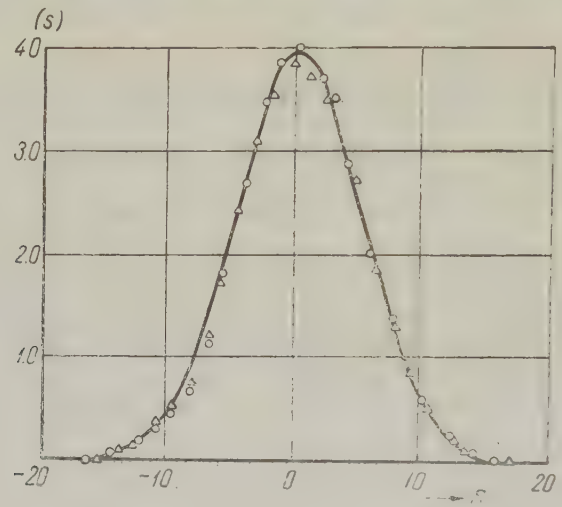
By most of the investigators lysozyme is considered as a polypeptide [1]. In order to gain a clear idea of the nature of lysozyme it is essential to determine its molecular weight. The power of lysozyme to diffuse through agar and cellophane membrane points to the small size of its particles. No direct determination was available, however, until done by Philpot on behalf of Abraham and Robinson [2]; with the aid of an ultracentrifuge he found that lysozyme has a molecular weight of the order of 18000. But this was no more than a provisional estimate, as pointed out by the two authors themselves [2].

The present authors undertook to determine the molecular weight of lysozyme by Lamm and Polson's diffusion method [3–5]. According to a special check-up by Polson [4], the results thus obtained were found to be in fair agreement with the data obtained by means of the ultracentrifuging method.

Lamm's method consisted in photographing the micrometrical scale through a column of liquid in which diffusion takes place. Owing to the vertical concentration gradient, the position of the lines of the scale on the photograph does not coincide with the control photograph of the scale, taken through a pure solvent. The differences between the positions of the line as recorded by the two photographs are measured with the aid of a microcomparator with an accuracy up to 2—3 μ, and serves to evaluate the concentration gradient of the substance diffusing into the solution, which in turn makes it possible to determine the coefficient of diffusion of this substance. At the same time a comparison of the normalized experimental curve with Gauss' theoretical curve gives an idea of the molecular homogeneity of the substance, unless some foreign anomalies make the picture obscure.

In order to find the molecular weight M it is necessary also to determine the relation between the semi-axes of the particles, b/a; this magnitude is obtained from the viscosity of diluted solutions of the substance investigated. The procedure, as well as the measuring device have been described in an other paper ([3]). Our photographs were taken with an objective of long focal distance, $F = 80$ cm.

The lysozyme preparation was made with fresh white of egg according to Borde's modified method, and was purified by repeated precipitation from 80 per cent alcohol according to Buyanovskaya ([1])*. A 0.5 per cent solution of the purified lysozyme preparation was next thoroughly dialysed at a refrigerator through a collodion membrane against distilled water; after this the solution was added sodium chloride (up to 1 per cent), and re-dialyzed against 1 per cent NaCl. The external 1 per cent NaCl solution was then used as the upper liquid for diffusion



——— Theoretical curve; ○Δ, experimental points

experiments. The external solution, containing 0.4 per cent of lysozyme + 1 per cent NaCl, pH = 6.34, measured with the aid of a glass electrode, served as a basic solution for the experiments on diffusion and viscosity. The biological activity of the solution was maintained normal, according to Buyanovskaya. The experiments were run at a temperature of 20°C ± 0.01. The values of the diffusion coefficient $D \cdot 10^{-7}$ within a range of 7200 to 10800 seconds were found to be 8.8; 8.4; 8.7; which makes a mean $D_{20,w} = 8.6 \cdot 10$ cm²/sec. The good agreement between the normalized experimental data and Gauss' theoretical curve (see figure) testifies to the molecular homogeneity of the preparation and to the absence of anomalies due to the charge of the particles, or the concentration of the solution. The value of the viscosity coefficient of the lysozyme solutions diluted to below 0.1 per cent, in 1 per cent NaCl, is

$$K = \frac{\eta_{sp} - 1}{vc} = 15.8$$

where η_{sp} is the specific viscosity of the solution; $v = 0.75$, the specific partial volume (in cm³/g); and c, the concentration of lysozyme in g/cm³. According to Polson's equation ([4]), $K = 4.0 + 0.098\,(b/a)^2$, whence with $K = 15.8$, $b/a = 11$; this value of b/a corresponds to the asymmetry coefficient ([5]) $f/f_0 = 1.58$.

From the figures cited above it is possible to calculate the molecular weight M according to formula ([4,5]): $M = \frac{k'}{(f/f_0)^3 D^3 s}$, where $k_{20,w} = 2.41 \cdot 10^{-14}$; substituting $f/f_0 = 1.58$, $D_{20,w} = 8.6 \cdot 10^{-7}$ and $v = 0.75$, we obtain $M = 12800$.

* The authors' sincere thanks are due to Professor Z. Ermolieva and Dr. Buyanovskaya for having supplied the authors with a lysozyme preparation and shown constant interest in their work.

Thus, according to our determinations, lysozyme has an approximate molecular weight $M=13000$, and consists of monodisperse elongated particles, showing a relation between their semi-axes, $b/a=11$.

It is of interest to note that the elementary composition of lysozyme, as determined by Meyer and co-workers ([6]) (*viz.* C—48.65 per cent, H—6.44 per cent, N—15.33 per cent, P—0.25 per cent, ash—3.31 per cent, and S—0.64 per cent), implies an easily computable empiric formula $C_{500}H_{790}O_{195}N_{135}S_{2\cdot5}P$, corresponding to a molecular weight of 11920. In order to obtain an integral number for the ratio S : P, the authors cited above doubled the formula and obtained $M=24000$; however, the accuracy of the determination of S seems hardly to justify this procedure. Putting S:P=2 or 3, we have an approximate figure of $M=12\,000$, which is in fair agreement with the figure obtained by us ($M=13\,000$).

Lysozyme is usually supposed to be an enzyme of a mycolytic action. However, if its low molecular weight be taken into account, the effect it produces upon the walls of the bacterial cells may be assumed to be analogous to the solvent and peptic effect produced by low molecular fractions upon higher molecular products, which has been described in colloid chemistry for certain colloids: serum-globuline, gelatine, cellulose, esters and other high molecular substances. The relatively low specificity of lysozyme, the rapidity of its effect, its capacity to be adsorbed by sensitive cells, the swelling of the walls of bacterial cells before lysis, the existence of a direct proportionality between the lytic effect of lysozyme and its concentration, all these principal facts seem to be in accordance with the above assumption. It is of interest to note that alcohol treatment continued during several hours and treatment with chromic acid continued during 10 minutes, having a tanning effect upon the membranes of bacterial cells, make the bacteria resistant to the effect of lysozyme.

Such facts as the existence of a pH optimum, the variation taking place in the electrokinetic potential of the bacteria in the presence of lysozyme, the influence of salts upon the effect of lysozyme and many others seem to shed light on the rôle played by polar groups in the adsorption of lysozyme. The lytic effect of lysozyme, connected with its low-molecular nature, is thus likely to be nothing else than a special case of the solvent action of low molecular fractions, which is well known in colloid chemistry.

Laboratory of Colloid Chemistry.
All-Union Institute for Experimental Medicine.
Institute of Biochemistry.
Academy of Sciences of the USSR

Received
8. XII. 1944.

REFERENCES

[1] З. Ермольева, Усп. совр. биол., **9**, 68 (1938). [2] E. Abraham and R. Robinson, Nature, **140**, 24 (1937); Biochem. J., **33**, 622 (1939). [3] O. Lamm and A. Polson, Biochem. J., **30**, 538 (1936). [4] A. Polson, Koll. Z., **87**, 149 (1939). [5] А. Пасынский, Усп. хим., **10**, 526 (1941). [6] K. Meyer, J. Palmer, R. Thompson and D. Khorazo, J. Biol. Chem., **113**, 479 (1936).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

GEOLOGY

QUATERNARY GLACIATION OF THE SUNNAGHIN RANGE (SOUTH YAKUTIA)

By E. V. PAVLOVSKY and G. K. SEMIGUZOV

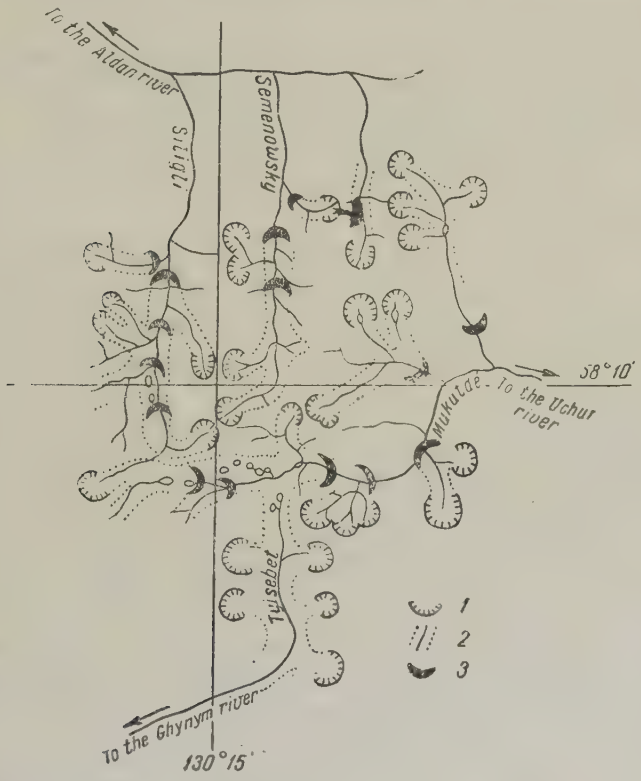
(Communicated by V. A. Obruchev, Member of the Academy, 29.XII.1944)

The South of Yakutia (the area of the Aldan plate) has been the subject of intense study during the recent years. Instead of the former conception of the Aldan plateau a new one has been advanced, implying the existence of a great arch uplift stretching from the middle course of the Olekma River westward to the lower course of the Uchur River in the East. This Aldan-Uchur arch uplift is well reproduced by the relief. In certain places its highest points reach 2000 m of absolute height. Along its northern foot flows the Aldan in its middle portion. Parallel to the southern limb of the arch the latitudinal parts of the upper course of the Aldan and Timpton rivers are located, and flows the Ghynym River, a large left-hand tributary of the Uchur. The meridional parts of the valleys of the Aldan and Timpton rivers, the lower course of the Uchur cut the arch uplift across its strike, abounding in shallows and rapids. The width of the arch uplift markedly decreases as it proceeds from the Olekma (width of the arch 200 km) eastward to the Uchur (width of the arch 50 km). On the right bank of the Uchur the arch uplift plunges like a huge anticline, merging with the flat relief of the Aldan-Lena plateau. The details of the structure of this arch and the age of this remarkable structure will not be discussed here. This interesting problem will be the subject of a special paper. Let us point out in this place that the Aldan-Uchur arch uplift is directly related to a still larger structure known by the name of the Baikal arch (²). The Aldan-Uchur arch uplift is the extreme western periclinal part of the Baikal arch, with which the existence and development of the large depressions of the Baikal plateau is related.

The various areas of the Aldan-Uchur arch have names of their own. Thus, its eastern part is frequently named the Sunnaghin Range. This range was studied by the author in the summer of 1944 in connexion with a search after new deposits of phlogopite. In the area of the divide of the Siligli River (right-hand tributary of the Aldan), the Muguskan-Uchur and Muguskan-Ghynym, corresponding to the area of the highest points of the arch, the maximum marks reach 1850 m absolute. The relief of the divide is alpine, closely resembling—even to a complete identity—the landscape of the high-mountainous areas of the other portions of the Baikal arch—the Primorsky (Coastal) Range and the Udokan, Kodar, Kalar and other ranges (²). The country is dissected into a number of «goletz» (bare mountains) and «goletz» groups. On the summits of the «goletz» terraces remnants of an ancient peneplain are not infrequently preserved. The shape of the «goletz» is variable—pyramidal, conical, *etc.* The slopes of the «goletz» are gneised by numerous cirques and corries. On the steep walls of the corries many-year old firn basins lie in

detached patches (absolute height 1375 m). The bottoms of the corries are in a number of cases filled with rounded or oval lakes, dammed by a barrier either of country rocks or of heaps of morainal debris. The mouths of the corries are everywhere located on the same absolute height of 1250—1300 m. The corrie areas at present play the rôle of drainage basins for the recent river valleys.

In the upper parts of the main river valleys—of the Siligli and Mukumde rivers—the corries are grouped in a whole system, being arranged in the shape of an amphitheatre. From each corrie down the valley stretches a glacial trough with an ideal U-shaped transverse profile. The bottom of the glacial trough is usually filled with ground and marginal moraines, composed mainly of boulders and dlocks of Archaean crystalline schists and granites. A lesser part in the composition of the moraines is played by yellowish-brown loam. The boulder material is not



Geomorphological scheme of a part of the Sunnaghin Range, Ssale 1 : 200 000. Legend; *1.* eirques and corries; *2.* glacial troughs; *3.* ridges of terminal morainesand recessional moraines

graded with regard to size. The ground and marginal moraines form on the bottom of the glacial trough systems of hills oval in shape, stretched along the valley and separated by small troughs without drainage, usually swamped. Some troughs are fulled with shallow lakes. On the bottom and at the mouths of the corries, on the slopes of the corries, occasionally on their bottom amidst moraines stand out *roches moutonnées*, carved by ice, with the steep slope pointing down their valley, and their gentle slope pointing upwards. In some places the smooth descent of the bottom of the trough valley is broken by rather low (5—15 m) rock bars.

From the corrie regions down the trough valley the ground and lateral moraines are traceable for several kilometres. Thus, along the local valleys—the Siligli, Mukumde, in the upper course of which the corries are the most extensively developed, the moraines stretch along the bottom of the glacial trough for 8 km from the mouth of the main corrie. In other valleys this figure is considerably smaller, constituting 2 to 4 km. In the lower end of the morainic tongue, both in the main and in the minor valleys, are located terminal morainic ridges barring the valley. These

are composed of a boulder material enclosed in yellowish-brown loams, similar to that of the ground moraines. It is of interest to note that in all the large valleys, where the development of the morainic complex reaches its maximum intensity, the absolute mark of the terminal morainic ridge does never descend below 1000 m. In the valleys of the Siligli and Mukumde above the sandy morainic ridges there are still left traces of ridges of recessional moraines, barring the valleys athwart.

The whole complex of glacial deposits and formations is fresh; it has only to a small extent suffered destruction during the post-glacial time. The destruction processes were manifest in an appreciable weathering of the outcrops of country rocks at the highest points of the relief, in the walls of corries and glacial troughs, and in the bottom part of the latter. Owing to this fact, the lateral and ground moraines filling the lower part of the trough valley are in some places encumbered with coarse-blocky sharp-angular stony talus. A part of the *roches moutonnées*, while retaining their shape, consist not of a monolithic country rock, but of a system of blocks, retaining their spatial arrangement but separated by deep gaping crevasses. The work of post-glacial erosion has brought about but insignificant results. The beds of rivers and springs have been incised 1.5—2 m into the moraines, having locally cut a trench-shaped bed in the country rocks. The coarse sandy-boulder post-glacial alluvium, formed chiefly at the expense of a washing of the moraines, composes a rather low (1—1.5 m) single terrace leaning against the glacial deposits.

The system of glacial troughs is almost entirely occupied by elements of the present river network. In some places only, such as, *e. g.*, the summit of the Tyisebet, the bottom of the open trough valley plays the part of a divide between the basins of the Uchur and Ghynym rivers.

The bulk of observations made within the area of Aldan-Uchur divide point to the existence at the highest points of the Aldan-Uchur arch uplift of corries and small valley glaciers; during the recent geological past, the post-Pliocene, they have left ineffaceable traces on the whole relief of the country. Analogous phenomena should evidently have taken place in all other areas of the latitudinal arch, which have an identical absolute height with that of the area studied by the author. If we are to judge by its type and age, the glaciation of the Aldan-Uchur divide seems to be synchronous with the last glaciation whose traces are so strikingly pronounced west of the Aldan region in the Olekma-Vitim mountainous country and in the near-Baikal area ([1], [2]).

New observations went fully to confirm the views of Obruchev concerning the distribution of traces of glaciation in the east of Siberia, expressed by him in a number of papers, in particular in the well-known review of the evidence of the glacial period in Northern and Central Asia ([1]).

Received
29. XII. 1944.

REFERENCES

[1] В. А. Обручев, Бюлл. Четвертичной комиссии АН СССР, № 3 (1931). [2] E. V. Pavlovsky, Bull. Acad. Sci. URSS, sér. géol., No. 2 (1937).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

GEOLOGY

NEW DATA ON THE OCCURRENCE OF ORES IN THE INTRUSION OF THE LITTLE CAUCASUS (AZERBAIJAN SSR)

By A. N. SOLOVKIN

(Communicated by D. S. Beliankin, Member of the Academy, 29.IV.1945)

A number of large intrusive massifs of the eastern part of the Little Caucasus was studied by the author the last few years (1940—1943). Many of them had neither been mapped nor described. Neither was there any information as to the occurrence of ores in them.

The most interesting and one of the largest massifs in the Little Caucasus is the intrusive massif named by the author (1) the Dalidag intrusive. It is located in the upper course of the Aker and Terter Rivers and occupies an area of over 120 km² of the basins of these rivers.

The Dalidag intrusive is a typical discordant pluton, breaking through a series of deposits from the Turonian to the Oligocene and overlain only by Quaternary lavas, pebblestones and loess-like loams. It refers to the Oligocene intrusion extensively developed in the Little Caucasus (2,3).

A very important and noteworthy fact is that this pluton has been exposed by erosion from the apical portions down to depths of 2000 m. The altitude marks of the elements of the relief, built up on a substratum of intrusive rocks, range from 1500 to 3500 m. As a result, the intrusive rocks are in contact with diverse wall rocks. The chief among the latter are limestones, tuff-sandstones, porphyrites and tuff-breccias of a porphyritic composition.

The petrographical composition of the Dalidag pluton is a complex one. It consists of pyroxene granodiorite, quartz diorite, diorite, syenite-diorite, gabbro-diorite and porphyry-like varieties of these types of rocks. The syenite-diorite and quartz diorite are distinctly confined to the peripheral parts of the massif; the more basic types of rocks form schlieren. The porphyry-like varieties are characteristic of the near-apical portions of the massif. Aplites are of common occurrence. The general habit of the rocks is rather leucocratic. On the whole, the pluton offers a picture of an intense differentiation that had taken place in the past.

Considerable interest is attached to the area of contact- and hydrothermally mineralized rocks, associated with the massif, which covers about 250 km².

Within this area zones of skarned limestones, secondary quartzites, hornstones and ochreous hydrothermally altered rocks are recorded.

Particular importance is assumed by the near-contact areas of the massif, characterized by numerous large veins of ochreous quartz, attaining several hundreds of metres in length and 2—5 m in thickness. The total number of the veins recorded exceeds 100. Among the quartz diorites fine dissemination of molybdenite has occasionally been noted, accompanied by chalcopyrite and pyrite.

An important, although indirect evidence of ore occurrence in the given intrusive is the discovery of traces of gold and scheelite grains*, along with the common heavy minerals, in the black sands of the river system eroding the massif. It should be emphasized that showings of tungsten mineralization in this particular part of the Little Caucasus have been established for the first time, and they become the more important as the general geological setting of the region is such as to leave no doubt concerning the success of further search.

Finally, hydrothermal phenomena associated with the pluton are not infrequent; they are manifest in the formation of quartz-sulphide veins of polymetallic type. These veins form the Bashtibel deposit of zinc-lead-copper ores, as yet very little investigated; similar veins have been recorded from the region of the Chovdar village (kishlak) and near the Shikeran (Shakhkeran) village, where they are confined to vast zones.

From what has been said above it may be seen that the region of the occurrence of the Dalidag pluton is one of the most promising areas of the Little Caucasus, with regard to polymetals, tungsten, molybdenum and uranium ores. Its access involves considerable difficulties, but these are overbalanced by a complex of highly favourable geological factors, owing to which this region stands apart among the other regions of the Little Caucasus.

Received
29. IV. 1945.

REFERENCES

[1] А. Н. Соловкин, Труды АзГУ, II (1942). [2] К. Н. Паффенгольц, Изв. Геолкома, XLVIII, 3 (1929). [3] А. Н. Соловкин, Изв. АзФАН, № 1 (1940).

* The investigation of the black sands has been carried out by M. A. Zhitkova of the Institute of Geological Sciences of the Academy of Sciences of the USSR, and by the present author.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

MINERALOGY

ON THE FORMATION OF COLUMNAR CALCITE

By A. M. KUZNETZOV and D. P. PROCHUKHAN

(Communicated by W. I. Vernadsky, Member of the Academy, 11.XII.1944)

In the Upper Kungurian clay-marly facies of the western part of the Near Ural—Vereyno (1), Verkhne-Chussovskie Gorodki, the Shakva River, *etc.*—genetically interesting streaks of columnar calcite have been discovered.

The calcite streaks, from 1 to 25 cm thick, lie parallel to the beds in a 10-m series of clays and marls, stretching for several hundreds metres. The number of streaks is 1 to 7. The contact of calcite with the enclosing rock is distinctly pronounced: the boundary surfaces of the streaks look as if they were polished.

The dark-grey clay-marly rock containing the columnar calcite displays microscopic stratification and contains a profusion of minute particles of organic matter, numerous thin streaks of fibrous gypsum (gypsum occurs in the rock also in disseminated crystals), impregnations of iron oxides and pyrite crystals. In the near-contact zone the rock shows no stratification, but contains accumulation of pyrite and is almost devoid of gypsum.

Externally the columnar calcite resembles fossilized wood, and consists of columnar crystals of varying thickness, grown together to form a monolithic rock. The crystals are parallel to one another and their long axes are perpendicular to the contact with the enclosing rock; they form rhombohedra stretched in the direction of the edge, owing to which they resemble a trihedral prism. The crystals are parted along the faces of concrescence, and break along the perfect cleavage planes pointing to their long axis at an angle of 45°. Between the prisms there are not infrequently vertical hollow channels with a section up to 1 mm and 5—7 cm long. The width of the faces of the prisms ranges from 1 to 5 mm. There are varieties with a capillary section of the crystals.

The layers of columnar calcite have local breaks in the growth of the crystals expressed in the form of seams 0.5—1.5 mm thick. The seams are filled with finely crystalline tabular calcite, with disseminated particles of organic matter. The number of seams in the layer is 2—3, rarely over, but does not exceed 7. The terminations of the thin crystals are inclined (curved) in a direction opposite to that of the incline of the crystalline individuals, located on the other side of the seam.

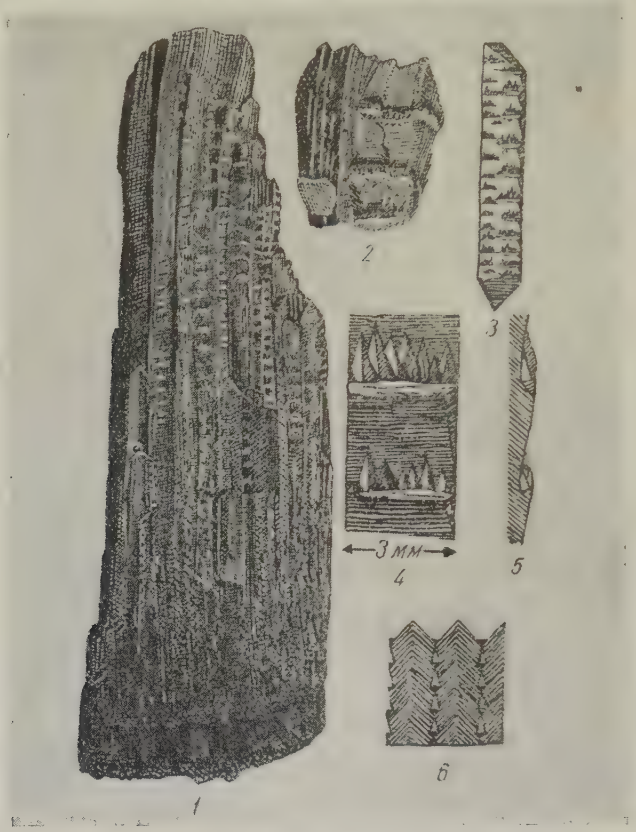
The colour of the mineral varies from light to yellow, pink, bluish to brown and dark-grey; it depends on the quantity and dispersion of the solid admixtures which are readily separated when calcite is dissolved in hydrochloric acid. The light varieties contain a negligible amount of impurities producing only opalescence in the solution, while the dark ones are enriched with insoluble material up to 12 per cent.

The light varieties are always composed of thin columns; similar columns are characteristic also of the dark varieties of calcite, if the insoluble material is

highly disperse. Such varieties are endowed with considerable strength. Comparatively coarse and polydisperse material is associated with coarsely columnar brown calcite.

The prisms of the brown variety, showing a width of faces over 1 mm, carry on their lateral surface very remarkable «bridges» of clastic material. These bridges are placed on slanting benches on the faces of the prism across the long axis of the crystal, at a distance equal to the width of the face, or two or three times as great.

The position of the bridges in space is like a spiral engirdling the prism, the



1, external appearance of calcite, specimen No. 3, 1 : 2;*2, part of specimen, natural size; *3*, prism, ×1.3; *4*, face of prism with bridges and crests, magnified; *5*, structure of face of prism, lateral view; *6*, scheme of growth of crystals and arrangement of bridges

space between two subsequent bridges being equal to the width of the face, or half of it. The length of the bridges, as well as their thickness, is proportional to the width of the faces; the length ranges within 1—5 mm, and the thickness, 0.1—2 mm.

Over the bridges small pyramidal vugs are located, the base of which is covered with a brush of crystals of calcite, with their acute extremities pointing upwards. The accompanying drawings illustrate the aforesaid.

The calcite columns in the transverse section are formed of lobate, uniaxial, optically negative crystals (size 0.15—0.40 mm), which in the longitudinal section give a peculiar cone-in-cone pattern, similar to the structure of calcite from series of different age (²), characterizing their conical growth in layers. The coefficient of refraction $N_g = 1.65$; $N_p = 1.48$; birefringence 0.17. Without analyser the crystals look colourless; with crossed nicols, faintly yellowish. Cleavage is perfect; hardness, 3—4; specific gravity, 2.66—2.76; not coloured with cobalt nitrate.

The mineral consists of $CaCO_3$ and a dispersoid evenly distributed throughout the mass of crystalline calcium carbonate, which gives a key for the understanding

of the mechanism of its formation. The size of the visible clastogenic particles measures 0.2—5 μ, less frequently 30—40 μ, and very seldom more; the mineral may be referred to crystal sols (3). The composition of the calcite is illustrated by the results of an analysis of three varieties.

Constituents	1, pink	2, dark-grey	3, brown
$CaCO_3$	100.0	88.05	89.41
$CaSO_4 2H_2O$	—	0.37	—
Dispersoid	traces	11,58	10.59
Specific gravity . .	2.71	2.76	2.66

The disperse silt (specimen 2) consists of colloidal silica, clay, iron sulphide, numerous spores and pollen (up to 200 specimens 50—70 μ in size in the preparation of specimen 3)*, well preserved, and black particles and films with a coaly lustre. Its general composition is: organic matter, 18.0; SiO_2, 56.8; Al_2O_3, 7.3; Fe_2O_3, 6.9; FeS, 5.5; MgO, 5.0 per cent.

From the aforesaid it follows that the combination of calcite with the dispersoid pyrite and organic matter, as well as the reverse ratio in the content of pyrite and gypsum of the enclosing rock, against a background of a periodic change of the sulphate sediments of the Permian (4), are evidence of the biogenic-reducing medium at the period of formation of columnar $CaCO_3$ from $CaSO_4$. The high solubility of calcium sulphate (5) as compared to its carbonate caused a constant oversaturation of the solution with $CaCO_3$ and its crystallization in the shape of columnar prisms with a one-sided supply of $Ca^{\cdot\cdot}$ and CO_3'' ions from the zone of reduction.

The rate of growth of the prisms was determined by the intensity of the biogenic process, and the latter, by the rate of the supply of the near-bottom region with sulphate and organic matter, where calcite columns were growing, fixing the settling suspension. The uniform distribution of the dispersoid in the calcite shows a constant and insignificant concentration (0.2 g/l) of the suspensoid. The formation of $CaCO_3$ probably occurred in a sufficiently deep part of the basin, or at periods when the erosion of land was at the point of extinction.

Assuming that the solution is saturated with calcium sulphate (2.08 g/l) and completely passes through biogenic reduction to carbonate, we see that 1 cm^3 of $CaCO_3$ corresponds to 1.6 l of the solution located over it at a height of 1.6 m. The particles of the dispersoid, with an average diameter of 1 μ, take 21 days to pass this way; within the same time a layer of calcite 1 cm thick is formed. On the average, 10 cm of calcite is formed during three quarters of a year. The formation of calcite layers probably proceeded at a sufficiently high rate, being associated with seasonal rhythms, and did not take centuries.

The arrangement of the bridges shows that the trihedral prisms of calcite growing formed concrescences of six individuals with a common periodically shifting cone of growth. The dispersoid settled upon the oblique microscopically uneven faces of the tops of the hexahedra. As the growth of prisms in layers proceeded at a high rate, the dispersoid was uniformly crystallizing; when the rate of growth was slow, part of the clastogenic material, particularly the coarse one, was sliding down the inclined plane to the intersection of faces and accumulated in the biheral corners, forming bridges. Consequently, light varieties of $CaCO_3$ were formed in pure solutions, while with high concentrations of the suspension it was clay limestones and calcareous marls. These interrupted the formation of columnar calcite and fixed a certain quantity of gypsum which was the larger, the lower the intensity of the biogenic reduction process limited by a supply of organic matter. In this way it is also possible to account for the reverse ratio in the content of pyrite and gypsum in the clay-carbonate complex of the rocks. A rhythmic accumulation of the dispersoid in the form of bridges, subsequently coated with growing

* The determination was kindly made by M. A. Sedova.

crests of calcite and overlapping faces with the prism tending to restore its lost regular form, characterizes the alteration of the cones of growth along the pyramidal-polyhedral surface.

The layers of columnar calcite were thus the result of a periodically arising biogenic process which went on under conditions of a supply of comparatively pure solutions of calcium sulphate, organic matter—the source of energy of the biogenic process, during the periods of a sharply reduced supply of clastic materials. Geographically, this took place at the boundary between the sulphate and the terrigenous facies. Therefore columnar calcite may serve as a correlative in the stratigraphic subdivision of the unfossiliferous Kungurian formation, which has still remained but poorly investigated.

Received
24. XI. 1944.

REFERENCES

[1] К. Матвеев, Тр. С.-Петерб. об-ва естествоиспыт., XLI, в. 1, 5—6 (1910). [2] W. H. Twenhofel, Treatise on Sedimentation, Baltimore, 1932. [3] Ф. В. Чухров, Коллоиды в земной коре, изд. АН СССР, 1936. [4] Н. П. Герасимов, Геолог. строение вост. нефт. обл., изд. АН СССР, 1940. [5] А. Ф. Капустинский, Произведение растворимости, ЖПХ, XVI, 1—2 (1943).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

MINERALOGY

A ZEOLITE FROM THE MORDENITE GROUP IN THE UPPER CRETACEOUS AND PALAEOGENE MARINE DEPOSITS OF THE EASTERN SLOPE OF THE URALS

By N. V. RENNGARTEN

(Communicated by D. S. Beliankin, Member of the Academy, 26.X.1944)

The marine sediments of the Danian stage and Palaeogene of the eastern slope of the Urals contain authigenic monoclinic zeolite from the mordenite group. In spite of the fact that this mineral is often contained in the rocks in considerable quantities, and that its distribution is of regional character, it is for the first time that the existence of zeolites, genetically related to the sea, has been recorded in the Urals.

Zeolite from the mordenite group has been discovered by the author in a number of places on the eastern slope of the Urals. It may be said that this mineral is a reliable correlative for the Danian stage, distinguishing the Danian strata from the lower horizons of the Cretaceous. Zeolite-bearing rocks have been encountered amidst Danian deposits of the Alapaev, Troitzk-Bainov, Kamensky, Sinara and Techen regions. The zeolite crystals are usually confined to sandy deposits and are found in fine (<0.01 mm) fractions of the sediments. In some specimens the zeolites make up about 70 per cent of these fractions (Table 1), and are present in the form of well-developed minute crystals and fine aggregates.

For the Palaeogene sediments zeolite is not so characteristic a marking mineral. It occurs in them but sporadically and, as a rule, in smaller amounts than in the Danian strata. However, among the Palaeogene deposits, too, there are sediments exceptionally rich in authigenic zeolites. Such sediments are the areno-argillaceous rocks enclosing the oxide manganese ores of the Polunochny deposit. Here zeolite is present not only as detached minute crystals and fine aggregates disseminated in the rock mass, but also in the form of large (up to 3—4 cm across) isolated accumulations. The latter in the form of white «eyes» stand out distinctly against the background of the grey clayey mass of the rock.

Table 1

Place of sampling	Granulometric composition, per cent				Per cent of zeolite in fractions <0.01 mm
	>0.25 mm	0.25—0.05 mm	0.05—0.01 mm	<0.01 mm	
The Sinara River	0.2	93.6	6.0	6.0	70
Sokolovsky bauxite deposit	1.8	93.9	0.9	3.4	60
Do	0.3	82.2	0.8	9.7	50
The Kolchedanka River .	—	76.5	1.1	24.0	70
Do	0.3	89.2	0.8	9.7	95

The rocks containing zeolite belong in most cases to fine-grained quartz sands. Their granulometric composition and the zeolite content in the clay fraction are given in Table 1.

The clastic grains are rounded, belonging mainly to quartz; glauconite, alkaline feldspars, siliceous aggregates are of less frequent occurrence. Zircon, rutile, tourmaline, garnet, glaucophane are present as accessory admixtures. Spicules of siliceous sponges are present. The fine fractions of these sands consist of well-formed small crystals of zeolite, lamellae of green chloride and clayey pelitomorphic material. Certain lamellae of chlorite contain small crystals of zeolite.

The zeolite crystals are on the average 0.05—0.01 mm in size. They are colourless and have a prismatic habit of distinctly pronounced crystallographic outlines. The crystals belong to the monoclinic syngony. The plane of symmetry parallel to the face of the second pinacoid [010], one double axis [010] and the centre of inversion were determined. The angles between the pinacoids are in general constant and equal: to 116 , the angle between [100] and [001], and to 125 , the angle between [001] and [101]. A perfect cleavage along [010] and a less perfect one along [001] are observed. The elongation of the mineral is negative; extinction approaches the straight one. The optical indicatrix is located in such a way that the N_p axis coincides with the third crystallographic axis; the N_m axis, with the second axis; and the N_g axis lies in the plane [010] at right angles to the first crystallographic axis. The following refractive indices of the mineral have been determined: along N_g, 1.487; along N_m, 1.485; along N_p, 1.483.

When taking samples of the mineral for chemical analysis, I did not succeed in completely separating zeolite from chlorite. The results of the analysis were found to be affected by the presence of the latter. Therefore, prior to proceeding to the calculation of the formula of zeolite, it is necessary to exclude from the data of analysis the amount of oxides falling to the share of chlorite.

Assuming the formula of chlorite to be

$$3MgO \cdot 2SiO_2 \cdot 2H_2O + Fe_2O_3 \cdot Al_2O_3 \cdot SiO_2 \cdot 2H_2O$$

we have to substract from the data of analysis MgO and Fe_2O_3, and the SiO_2, Al_2O_3 and H_2O, corresponding to them according to the formula (Table 2).

Table 2

	Chemical analysis of zeolite + chlorite		Zeolite, molecular quantity
	Composition, per cent	Molecular quantity	
SiO_2	63.27	1.055	1.008
TiO_2	0.64	0.009	0.009
Al_2O_3	15.22	0.149	0.127
Fe_2O_3	3.53	0.022	—
CaO	2.77	0.049	0.049
MgO	1.39	0.035	—
K_2O	2.77	0.019	0.019
Na_2O	0.69	0.011	0.011
$-H_2O$	10.47	0.586	0.586
$+H_2O$	6.11	0.339	0.272
	105.86		

The formula

$$2(Na,K)_2O \cdot 3CaO \cdot 8Al_2O_3 \cdot 67SiO_2 \cdot 54H_2O$$

is deduced.

This formula closely approaches the mordenite known in the literature, which has the composition:

$$4Na_2O \cdot 4CaO \cdot 8Al_2O_3 \cdot 64SiO_2 \cdot 50H_2O\ (^1)$$

However, the zeolite here described, cannot be named mordenite proper, since it is distinguished from the latter by a different orientation of the optical indicatrix and by a somewhat higher index of refraction.

In mordenite the optic plane is perpendicular to [010]; the refractive indices are: along N_p, 1.471; along N_g, 1.476. In the zeolite here described the optic plane is parallel to [010], and all the refractive indices are higher than 1.476.

The question as to the genesis of the monoclinic zeolite from the Danian and Palaeogene deposits of the eastern slope of the Urals has thus remained unsettled.

This mineral doubtless occurs in rocks *in situ*, and has not been brought-in from outside together with other terrigenous material. Indeed, it would be difficult to imagine such conditions under which, along the well-graded sandy material, fine pelitic particles, belonging exclusively to chlorite and zeolite, would be brought into the basin and precipitated there.

The occurrence of authigenic zeolite in marine deposits alone is evidence of the fact that this mineral is genetically related to the sea, and its formation could not take place under conditions of a late diagenesis when the sediments were quite isolated from the sea.

It is scarcely probable, however, that the precipitation of zeolite crystals from sea water proceeded strictly simultaneously with the deposition of the terrigenous material. Indeed, at the time of the accumulation of terrigenous material, particularly of graded sandy material, the sea medium was too mobile for the finest crystals of zeolites to settle. It seems, therefore, more probable that the crystals of zeolite were formed and developed in a fresh sediment while it was still below sea level and was soaked with sea water. Here the aqueous medium was little mobile, and its chemical concentration strongly differed from that of the normal sea concentration, owing to the fact that processes of decomposition of the terrigenous material were at work in the sediment. The regular development of crystals with the formation of terminal faces points to conditions favourable to a free growth of these crystals (the process of mineral formation was going on in a loose, unsolidified sediment).

As a result of an interaction of the sea water and of the products of weathering of the alumino-silicate terrigenous material crystals of zeolite began to precipitate.

The zeolite genetically related to sedimentary formations, may, like glauconite ([2]), be called a sea mineral. It is formed in the sea basin in the course of processes of halmirolysis of aluminosilicate terrigenous material.

The principal sources of removal of aluminosilicate terrigenous material in most cases probably were effusive rocks of the type of plagioclase porphyrites, extensively developed along the whole eastern slope of the Urals. In the sands of the Danian stage, for instance, chlorite lamellae, containing small crystals of zeolite, are encountered. These zeolite-chlorite aggregates developed possibly at the place of fragments of the hyalopilitic groundmass of effusives, where glass was replaced by chlorite, and plagioclase microlites, by zeolite.

The extensive development of colloidal substances that are so characteristic of the Danian-Palaeogene seas, as well as the formation of mineral compounds rich in water (zeolite) must have been favoured by the comparatively low temperature of the sea basin.

Undoubtedly, the formation of zeolites in sedimentary deposits is a much less infrequent phenomenon than it is usually believed. Further investigations in this field will shed additional light upon the problems of the chemistry, origin, *etc.* of this highly interesting group of aluminosilicates.

Received
26. X. 1944.

REFERENCES

[1] A. N. Winchell, Elements of optical mineralogy, Part II, Descriptions of minerals. 1932. [2] В. О. Малышева, Главконит и главконитовые породы Европейской части СССР, 1930.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

PHYSIOLOGY

SKIN RESPIRATION IN *CARASSIUS CARASSIUS*

By T. I. PRIVOLNEV

(Communicated by I. I. Schmalhausen, Member of the Academy, 16.I.1945)

Alongside of gill respiration, some fishes are able to consume oxygen through their skin. Attempts are known to connect the duration of survival of fishes outside the water with the intensity of their skin respiration (1). The purpose of the present work was to study, first, the intensity of skin respiration in young *Carassius carassius*, and, secondly, the relation between skin respiration and oxygen concentration in the surrounding atmosphere. Warburg's method was adopted for the experiments.

Young crucians, about 3 months of age, and 1.0—1.5 g in weight, were placed into containers about 50 cm³ in volume. In the first place oxygen consumption was determined in young crucians while kept in water. The latter was then removed, and oxygen uptake was now determined in the experimental animals kept in the moist atmosphere of the containers. In the latter case only slight and rare twitches of the gill opercles could be recorded, instead of the regular movements observed in the fishes when kept in water. In our experiments, the crucians survived in the moist chamber for no less than 15 hours.

In the experiments on the relation between skin respiration and oxygen concentration in the surrounding atmosphere the partial pressure of oxygen was varied by filling the containers with air and hydrogen mixtures at a ratio of 1/2: 1/2; 1/4: 3/4; 1/8:7/8; 1/16: 15/16; the partial pressure of oxygen ranged within 75, 37, 18 and 9 mm Hg, respectively.

When kept in water, the fishes display the two modes of respiration. Therefore, when determining oxygen uptake in a crucian staying in water, we obtain the sum total of oxygen consumed. The estimation of respiration in the moist atmosphere gives the value of the oxygen consumed through the skin. Having obtained the two values, it is possible to calculate the intensity of skin respiration in per cent to the total oxygen consumption. The experiments were carried out at 20°C. The data obtained are given in Table 1.

Table 1

Respiration of *C. carassius* (gill and skin) in cm³ per 1 kg of Weight per Hour

	No. of specimens			
	2	3	4	5
O_2 uptake in water (total amount of O_2 consumed)	376.0	329.0	389.5	342.0
O_2 uptake in moist chamber (skin respiration)	90.2	86.5	87.5	96.1
Skin respiration in per cent to total respiration	24.0	26.3	22.5	28.1

It will be seen from Table 1 that the amount of oxygen consumed by young crucians is equal to about 25 per cent to the total amount of oxygen uptake. We have also tried to estimate the intensity of skin respiration in young carp and perch, but these attempts proved a failure, for the experimental fishes did not survive in the moist chamber and perished within the first hour of experiment. These data seem to show that skin respiration is a feature peculiar to *Carassius carassius* and possibly to some other fresh-water fishes only.

Skin respiration in young crucians proceeds in a way independent of the oxygen concentration in the surrounding medium, as will be seen from the data of Table 2

Table 2

Intensity of Skin Respiration under Different Partial Pressure of Oxygen

Partial pressure of O_2 in mm Hg	150	75	37	18	9
Oxygen consumption through skin in cm^3 per 1 kg of weight per hour	85.4	92.3	96.9	81.5	83.6

The change in partial oxygen pressure from 150 to 9 mm Hg does not materially influence the intensity of skin respiration. On the other hand, oxygen consumption through the gills is directly determined by the partial pressure of oxygen in the medium.

It will be seen from Fig. 1 that oxygen consumption in *Carassius carassius* is less dependent on partial pressure when oxygen concentration in the surrounding medium is low, and *vice versa*. These relations are opposite to those usually observed

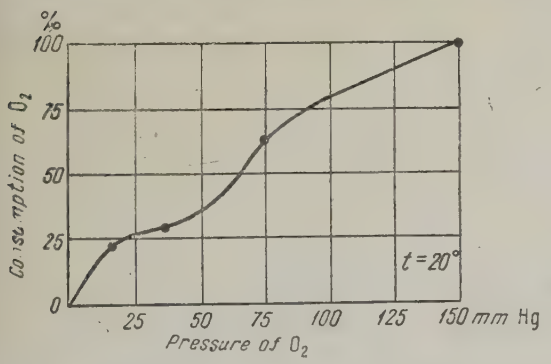


Fig. 1. Pressure of O_2 (in mm) plotted along the abscissa, consumption of O_2 (in per cent), along the ordinate

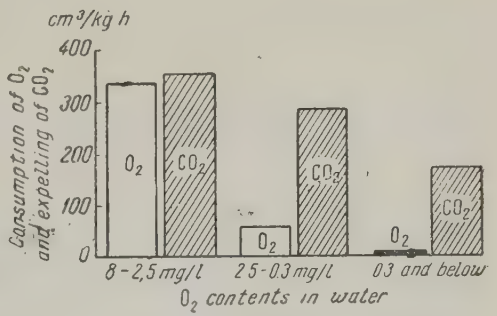


Fig. 2. Oxygen concentration in water (in mg per litre) plotted along the abscissa; oxygen uptake and carbon dioxide release (in cm^3 per 1 kg of weight per hour), along the ordinate

in fish: oxygen uptake is less dependent on partial pressure in the surrounding medium, when oxygen concentration is high. The relation here described is but an apparent one, being due to the skin respiration of the crucian.

As mentioned above, skin respiration proceeds more or less independently of the partial oxygen pressure, while gill respiration is reduced, as the partial oxygen pressure is decreased. It is quite clear that the decrease in the partial oxygen pressure will reduce the total oxygen uptake more strongly in the case where gill respiration is predominant, *i. e.* within a range of higher oxygen concentrations.

There is another peculiarity in the respiration of *Carassius carassius* which should be mentioned here. It was found in the course of the experiments here described that the respiration coefficient is strongly increased if oxygen concentra-

tion of the medium is reduced, and a large part of the oxygen is consumed through the skin. When partial oxygen pressure reaches 18 mm, the respiration coefficient is equal to 1.5—2, showing that a large part of CO_2 is released without oxygen consumption (Fig. 2). This observation should be interpreted to mean that the process of aerobiotic glycolysis is taking place when the oxygen concentration in the surrounding medium is strongly reduced.

Consequently, a skin respiration independent of the partial oxygen pressure, as well as of aerobiotic glycolysis enables the crucian to survive at such low oxygen concentrations as are lethal for most of fresh-water fishes.

All-Union Institute of
Lake and River Fisheries.

Received
2. I. 1945.

REFERENCES

[1] O. Polimanti, Arch. Anath. und Physiol., I Abt. (1911).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945 Volume XLVIII, № 8

PHYSIOLOGY

PHAGOCYTIC ACTIVITY OF LEUCOCYTES AS AFFECTED BY SERUMS CONTAINING LEUCOCYTIC TREPHONES

By A. A. SHUTOVA

(Communicated by A. A. Zavarzin, Member of the Academy, 24.III.1945)

The study of phagocytosis has revealed its dependence upon two agents, *viz.* the activity of the cell and the influence of the medium where the process is taking place.

Some of the authors are inclined to correlate the phagocytic activity of blood leucocytes with the age and the degree of maturity of the cell ([14],[19]). Others report individual variations in phagocytic activity transmitted by way of heredity ([13]). The influence of the medium, as determined by physical-chemical agents, has also been studied in detail ([11], [12], [15], [17], [18]).

The present author undertook to investigate the phagocytic activity of leucocytes in horse serum, both normal and containing leucocytic trephones; the work appeared to be promising, since, according to studies by Carrel ([7]), leucocytes cultivated in serums produce substances whose effect is to stimulate growth and activity of other cells.

The application of serums containing leucocytic trephones (they will in the sequel be termed «trephone serums», for brevity sake) to the treatment of wounds showed them to increase the local reactivity of the tissues, and to stimulate the regeneration process ([3]).

The appearance in the serum of leucocytic trephones caused by the life activity of leucocytes was likely to be due to the changes taking place in its structure and immunobiological properties. In order to shed some light upon this side of the question, we arranged a series of experiments on the phagocytic activity of leucocytes exposed to the influence of a horse serum which had served during 48 and 72 hours as a medium for cultivating leucocytes, after the method described by Khrushchov ([3]), whose sections were kindly put at the author's disposal and served as material for the present work. As control use was made of a horse serum previously kept in a thermostat for a period equivalent to that of the experimental serum. Both serums were then centrifuged and served as media for comparative experiments on phagocytosis.

10 series of the serum were investigated, 6 of which had been taken after 72 hours, and 4, after 48 hours of cultivation. The experiments were carried out according to Wright's method ([2]) with a 24-hour culture of staphylococci and human leucocytes, invariably taken from the same individual. The experimental and the control serums were tested simultaneously, using always the same bacterial suspension, 1 cm^3 of which contained 1 000 000 000 microbes.

Phagocytosis was observed at different time intervals, ranging between 5, 10 and 15 minutes, except for experiment No. 5, which was run at 10, 20 and 30-minute intervals. This arrangement enabled us to obtain a more detailed picture of the phagocytosis process, and of the changes suffered by leucocytes in the course

of its development. The phagocytic indices which served to express the average number of microbes captured by an individual leucocyte, were calculated from two or three smears, for 100 leucocytes in each of them. Polynuclear leucocytes and monocytes were recorded. The results of the experiment are summarized in Table 1.

Table 1

Phagocytic Indices

No.	Serum	Duration of experiment		
		5 min.	10 min.	15 min.
1	Trephone	—	3.66	5.36
	Control	—	2.08	12.73
3	Trephone	2.26	5.23	3.32
	Control	2.05	2.68	6.15
4	Trephone	0.77	3.56	4.08
	Control	1.03	2.59	3.24
6	Trephone	3.05	5.07	7.72
	Control	1.38	1.83	2.24
7	Trephone	4.2	5.67	8.08
	Control	1.7	4.1	3.5
8	Trephone	3.37	7.9	8.5
	Control	2.36	5.97	5.13
9	Trephone	3.34	7.09	8.34
	Control	1.86	3.98	10.03
10	Trephone	8.31	11.7	8.59
	Control	5.75	14.14	11.28
11	Trephone	6.51	6.66	5.53
	Control	3.72	5.09	9.63
		10 min.	20 min.	30 min.
5	Trephone	2.75	3.56	8.82
	Control	1.03	2.21	11.17

The data contained in Table 1 permit of the following conclusions. In the five-minute tests the difference between the phagocytic indices obtained for the control and the trephone serums is found to vary in the degree of expression (see experiments 3 and 4). After 10 minutes this difference is found to have increased (especially in experiments 3, 6 and 9). By this time the two serums, the control and the trephone one, offer strikingly different morphological pictures. In the trephone serum the digestion of microbes proceeds at a much higher rate: along with stained microbes, number of semidigested, poorly stained forms become observable in the leucocytes; occasionally utterly discoloured microbes are encountered, which can be identified only by their low refraction index against the background of the protoplasm. These phenomena are still more conspicuous in the 15-minute experiments. The protoplasm of the leucocytes is strongly altered, vacuolized and shows scarcely any granulation. The phagocytic index undergoes a sharp drop in some of the experiments (1, 3, 9, 10, 11). In experiment 10 the decrease in the phagocytic index can be recorded as early as within the 10-minute interval. We consider these serums to be endowed with the highest activity. Under their influence a large number of microbes trapped by the leucocytes are nearly digested within 15 minutes, so as to make any count impracticable, which is the cause of the drop in the phagocytic index. These phenomena make us suppose that the rate of digestion is strongly increased in the trephone serum.

We cannot decide whether the stimulating effect is due to the production of special activators, or to some changes taking place in the substratum of the serum during the time of the cultivation of leucocytes; it is possible that both these agents are operative here. The only thing we wish to point out is the slight alteration of the proteinic component in the trephone serum, which had been observ-

ed in this laboratory. But even insignificant changes in the chemical structure may suffice to bring about important shifts in the picture of phagocytosis.

The phagocyting leucocytes are subject to considerable changes. They swell and increase in volume, become labile and are deformed. Their granular structure is frequently obliterated, giving place to a profuse formation of vacuoles in the protoplasm.

All forms of leucocytes are involved in the process of phagocytosis. Along with neutrophiles and monocytes, that had been described in classic literature, we happened to observe phagocyting eosinophiles which would not only capture but even digest the microbes; phagocyting basophiles and lymphocytes were also encountered. The changes suffered by the latter were especially striking. In the process of phagocytosis, the limphocytes rapidly completed their developmental cycle and attained the stage of macrophages. We observed a large number of transitional stages in the development of lymphocytes and their transformation into monocytoid forms, which became gigantic macrophages. These cells were occasionally found to be arranged individually or in pairs, while in other cases we observed accumulations of 5, 8, 10 cells and even more, densely packed together. The agglutination phenomenon was observed in some of the experiments, irrespective of the form of the leucocytes.

In ordinary lymphocytes scattered all over the preparation we happened to observe different developmental stages denoted by us as «transitional forms». They were characterized by nuclear changes involving both shape and colour. The following picture was of the most frequent occurrence. The nucleus of a small lymphocyte assumed irregular outlines, growing poor in chromatin and strongly increasing in size. The alteration in the nuclear apparatus was accompanied by a swelling of the protoplasm, owing to which the cell attained the size of a monocyte. Along with the proliferating types there could also be observed forms of amitotic division, in which nuclear division was not accompanied with division of the protoplasm. In none of the cases did we observe mitotic division.

When calculating the leucocytic formula, all lymphocytes, transitional and monocytoid forms and macrophages were taken into account, and their ratio to the total number of granulocytes was computed. The results of this computation is given in Table 2 which illustrates the changes suffered by the lymphocytes in the course of phagocytosis at 5, 10 and 15-minute intervals. The table summarizes the results of 8 experiments. The parallel control tests carried out under identical conditions, save for the microbes, also yielded negative results with respect to the development of macrophages.

As may be seen from Table 2, the development of macrophages cannot be observed in all experiments. It can be traced at a various degree in experiments 3, 4, 5, 6, 7, 8, while in experiments 9 and 10 it is altogether absent. It does not always follow the same course both in the control and the trephone serums. No growth of macrophages can be observed in the 5-minute tests, except for experiment 3; in 10-minute tests it attains its maximum expression, whereas occasionally it is not observed in 15-minute tests. The last circumstance may be accounted for by the lability of the cell, called forth by its physical-chemical state, and perhaps also by the imperfection of the technique. Swollen leucocytes are readily deformed and destroyed during phagocytosis. In our sections we observed many large, pale nuclei, which were frequently found to preserve traces of light-blue protoplasm; we cannot refer these nuclei to anything else than the nuclei of macrophages.

If we compare the increase in the size of macrophages in the two serums, the trephone serum does not compare favourably in this respect with the control one. In a number of experiments the macrophages do even develop at a lower rate in trephone serum. These results stand in need of a thorough check-up and further investigation.

The comparison of the results of 5-minute and 10-minute experiments and the observation of the picture of phagocytosis—appearance of numerous transitional forms between lymphocytes and monocytes—make us conclude that in this

Table 2

No. of exper.	CONTROL SERUM									TREPHONE SERUM								
	Duration of experiments																	
	5 minutes			10 minutes			15 minutes			5 minutes			10 minutes			15 minutes		
	Lymphocytes	Transitional	Macrophages	Lymphocytes	Transitional	Macrophages	Lymphocytes	Transitional	Macrophages	Lymphocytes	Transitional	Macrophages	Lymphocytes	Transitional	Macrophages	Lymphocytes	Transitional	Macrophages
3	68.9	12.5	18.6	69.2	12.6	18.2	75.2	12.5	12.3	77.2	12.8	10	64.5	16	19.5	80.8	10.8	8.4
4	75.2	10.4	14.4	71.7	7.2	21.1	73.9	10.8	15.3	82.2	9.7	8.1	63.7	15	21.3	72.6	5.6	21.8
6	71.7	15.3	13	60.6	19.6	19.8	84.6	11.1	4.3	60.5	23.9	15.6	72.6	19.5	7.9	76.6	12.9	10.5
7	90	5.7	4.3	46.4	5.7	47.9	81.2	12.3	6.5	91.8	5.5	2.7	70.8	9.4	19.8	71.1	3.2	26.4
8	81.7	4.8	13.5	79.5	8.3	12.2	80.2	11.4	8.4	91.6	7.5	0.9	68.6	7.2	24.2	92.7	5.3	2
9	96.2	2.3	1.5	87.8	9.1	3.1	86.2	8.4	5.4	87.7	2.9	9.4	81	5.8	13.2	85.5	8.2	6.3
10	64.8	23.1	12.1	63.4	28.2	8.4	80.6	16.6	2.8	84	12	4	77	21.3	1.7	61.9	26	12.1
	10 minutes			20 minutes			30 minutes			10 minutes			20 minutes			30 minutes		
5	63.8	10.1	25.6	62.0	21.5	16.5	81.5	11.1	7.4	67.8	16.4	15.8	80.6	12.6	6.8	75.4	19.8	4.8
	Experiments without microbes																	
	5 minutes			10 minutes			15 minutes			5 minutes			10 minutes			15 minutes		
6	78.8	12.4	8.8	66.7	20	13.3	91.9	1.9	6.2	85.1	6.5	8.6	74	11.2	14.8	92.4	3.8	3.8
8	—	—	—	72.3	15.5	12.2	—	—	—	—	—	—	73.8	16.7	9.5	—	—	—
9	—	—	—	77.4	9.1	13.5	—	—	—	—	—	—	82.9	10.4	6.7	—	—	—
13	76	13.3	10.7	82.5	10.4	7.1	83.6	7.5	8.9	82.9	10.6	6.5	82.9	12.4	4.7	79.9	7	13.1

case we have to do with a regular process of development and transformation of one definite cell form to another. This process is called forth by such a powerful agent as the phagocytosis.

The changes in the lymphocytes here described have previously been observed in tissue cultures by different workers, as, for example, by Caffier([6]) and Maximov([16]), as well as by Bloom([5]) and Conway([9]) who succeeded, by introducing in a living organism *Bacillus monocytogenes*, to produce in this organism hypermonocytosis and to observe the transformation of lymphocytes into monocytes.

It is also of interest to compare the results obtained by the present author with those of Carrel and Ebeling([8]) who found lymphocytes and monocytes able of developing and surviving in a serum, and demonstrated the dependence of the morphological structure of the cell on its metabolic state. Baker([4]) showed the important rôle played in the development of monocytes by the products of protein breakdown obtained through the action of enzymes. In the light of all these investigations, as well as of later evidence as to the rôle played by nucleic acids in the development and growth of the cell ([1],[10]), taking also account of the active rôle of leucocytes in nucleic exchange([20]), we are brought to believe that the changes observed in lymphocytes may well be due to the products of decomposition of the bacterial antigen and to alterations in the composition of the proteins of the serum itself, due to the enzymic activity of the leucocytes.

Institute of Cytology, Histology and Embryology.
Academy of Sciences of the USSR.

Received
24. III. 1945.

REFERENCES

[1] Б. Кедровский, Успехи совр. биол., **15**, в. 3 (1942). [2] А. Райт, Основы вакцинотерапии, 1908. [3] Г. Хрущов, Роль лейкоцитов крови в восстановительных процессах в тканях, изд. АН СССР, 1945. [4] L. Baker, Journ. exp. Med., 57 (1933). [5] W. Bloom, Fol. Haem., 37 (1928). [6] P. Caffier, Arch. exp. Zellforsch., 4 (1927). [7] A. Carrel, J. Am. Med. Assoc., **82**, 4 (1924). [8] A. Carrel and Ebeling, J. exp. Med., 38, 5 (1923). [9] E. Conway, Arch. Path., 25 (1938). [10] A. Fischer, Nature, 144 (1939). [11] H. Hamburger u. J. Haan, Bioch. Z., 24 (1910). [12] H. Hamburger u. E. Hekma, Bioch. Z., 7 (1907). [13] L. Hektoen, J. Am. Med. Assoc., 57 (1911). [14] A. Hertzog, Am. J. Path., 14 (1938). [15] R. Höber u. J. Kanai, Kl. Wochenschr., No. 5, 209 (1923). [16] A. Maximov, Arch. exp. Zellforsch., 5 (1928). [17] J. Ouweele n, Pflüg. Arch. f. d. Ges. Phys., 164 (1916). [18] E. Philipsborn, Kl. Wochenschr., No. 9, Jahrg. 5 (1926). [19] M. Strumia and F. Boerner, Am. J. Path., 13 (1937). [20] M. Tschernoruzki, Bioch. Z., 44 (1912).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

ENDOCRINOLOGY

ON THE INFLUENCE OF ESTROGENS UPON THE THYROTROPIC REACTION

By P. A. WUNDER

(Communicated by I. I. Schmalhausen, Member of the Academy, 1.I.1945)

When analysing the fundamental agents influencing the susceptibility of the organism to a given thyrotropic hormone, it is necessary to study the rôle of the functional state of the sex glands in determining the degree of the reactivity of the animal. As has been pointed out elsewhere(1), the animals most sensitive to the thyrotropic agent of the pituitary body are infantile guinea-pigs, chicks and winter sparrows (*Passer domesticus*)(2,3) *i.e.* animals with underdeveloped sex glands. It is thus natural to conclude that the hypofunction of the gonads increases the degree of the thyrotropic reaction of the organism, while an increased secretion of the sex hormone must suppress the action of the thyrotropic hormone. Our experiments carried out on sparrow males showed, however, that the stimulation of the gonad function during the winter by means of lengthening the light day has no influence upon the susceptibility of sparrows to the thyrotropic action of hypophysal preparations(1). The increased hormonal function of the testes thus fails to affect the thyrotropic reaction in sparrows. It was of interest to study whether a sex hormone introduced from the outside, estrogen in particular, will inhibit the thyrotropic reaction.

The experimental procedure adopted was as follows. 5-day-old chicks were used as a material for the investigation. They received daily one intramuscular injection of oil solution of estrogen during 10 days. From the sixth day onwards the estrogen injections were followed by subcutaneous injections of soda extract from bull pituitary bodies, which were intended to stimulate the thyrotropic reaction. The injections were made two times per day during 5 days. The chicks of the control group were subjected to a similar treatment, with the only difference that instead of the oil solution of estrogen they were given the same doses of sunflower oil. Benzoate-folliculin, α-folliculin, and stilbestrol were used as estrogen preparations*.

The chicks were sacrificed 5 days after the injection of the thyrotropic extract. Their thyroid glands and oviducts were weighed with the help of torsion-balance. The increase in the size of the oviducts was taken as an index of the activity of the estrogen introduced. The results obtained in the experiments with stilbestrol and benzoate-folliculin are given in the table.

It will be seen from the table that the injection of the benzoate-folliculin caused an increase in the oviducts. The weight of the oviducts in experimental chicks was nearly ten times as large as in the control group. The effect of estrogen upon the thyrotropic reaction was not always the same, the weight of the thyroid bodies in the experimental chicks being somewhat less than that of the control ones (see table). However, this difference, being on the average equal to 0.9—1.1 mg, is

* Our heartiest thanks are due to Professor J. Kabak and Dr. M. Lapiner for having kindly supplied us with stilbestrol and α-folliculin.

Degree of Thyrotropic Reaction in Chicks as Influenced by Stilbestrol and Benzoate-Folliculin

Estrogen	Total amount of estrogen (mouse units)	Total amount of hypophysis in mg	Number of chicks	Weight of oviducts				Weight of thyroid glands (in mg)			
								For whole group		For female group	
				number of oviducts	minimum, mg	maximum, mg	average, mg	number of glands	average weight, mg	number of glands	average weight, mg
Benzoate-folliculin .	1 cm³ (= 5000 m. u.)	30	13	10	37	90	63±5.6	13	6.6±0.38	10	6.9±0.43
1 cm³ of oil (control)	—	30	14	9	5	8.5	6.56± ±0.52	12	7.5±0.63	8	8±0.83
Stilbestrol	1 cm³ (= 30000 m. u.)	35	20	9	650	1040	906.6± ±69.75	20	9.2±0.5	9	9.07± ±0.76
1 cm³ of oil (control)	—	35	20	11	7.3	14.2	9.1±0.69	20	9.1±0.46	11	9.35± ±0.47

no significant one. The experiments with stilbestrol are especially interesting because of the high biological activity it shows in the organism of mammalians. Our experiments have shown this synthetic estrogen to have a strong influence upon the oviducts of the chicks. The average weight of the oviducts in the experi-

1—3, oviducts of chicks given hypophysal extract + 1.5 mg of stilbestrol; *4*, oviducts of chicks given hypophysal extract + 0.4 mg of stilbestrol; *5*, oviducts of chicks given hypophysal extract + 1 mg of α-folliculin; *6*, oviducts of chicks given hypophysal extract + 5000 m. u. of benzoate-folliculin; *7*, oviducis of chicks given hypophysal extract alone

mental chicks was equal to 906.6 mg, while that in the control chicks was equal to 9.1 mg only. Thus, a 100-fold increase in weight of the oviducts was obtained

under the influence of 1.5 mg estrogen introduced during 10 days. But in spite of this effect, the thyrotropic reaction of the chicks was similar in both groups under consideration. No inhibitory effect of stilbestrol on the thyrotropic reaction could be observed.

The experiments were repeated with smaller doses of stilbestrol (0.4 mg) and α-folliculin (1 mg). All the estrogens were found to produce a stimulating effect upon the oviducts. Stilbestrol showed the highest activity. 0.4 mg of this estrogen caused the oviduct to increase more than did 1 mg of folliculin, the weight of the oviducts being equal to 445 and 117 mg, respectively.

According to our data, stilbestrol showed the strongest influence upon chick oviducts. Its activity was found to be nearly as high as that of ovarian sex hormone (see figure). However, these experiments also failed to demonstrate any decrease in the sensitivity of the organism to the thyrotropic stimulant of the pituitary body.

Thus, we are brought to the conclusion that estrogen cannot be regarded as an agent inhibiting the action of the thyrotropic hormone, in spite of the fact that this opinion based on some experiments with mammalians ([4],[5],[7]) is rather spread in literature. Therefore, the high susceptibility of chicks, winter sparrows, and infantile guinea-pigs is not directly connected with the hypofunction of the gonads and with a poor secretion of the sex hormone, since neither endogenic nor exogenic sex hormones are found to inhibit the action of the thyrotropic preparation.

Another conclusion as to the estimation of the treatment of Basedow's disease by prolonged administering of folliculin is to be made from the data presented above. Cases are known in literature where recovery from Basedow's disease took place after estrogen therapy ([6]). According to our data, these cases should be interpreted in the sense that recovery took place owing to the decrease in the secretion of the thyrostimulant from the pituitary body, leading to the restoration of a normal function of the thyroid gland, rather than to the direct influence of the sex hormone and inhibitory effect of the thyrotropic hormone on the gland itself.

Saratov State University.

Received
1. I. 1945.

REFERENCES

[1] P. Wunder, C. R. Acad. Sci. URSS, XXVIII, 361 (1940). [2] P. Wunder and K. Wibe, ibid., XXVIII, 358 (1940). [3] П. Вундер и К. Вибе, Учен. зап. СГУ, **15**, 165 (1941). [4] С. Гурылева, Бюлл. экспер. биол. и медицины, **4**, 205 (1937). [5] M. Aron et J. Benoit, C. R. Soc. de Biol., **109**, 923 (1932). [6] J. Lederer, Revue Belge Science Médic., **11**, 326 (1939). [7] J. Rowlands and A. Parker, Biochem. J., **28**, 1829 (1934).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

EXPERIMENTAL MORPHOLOGY

ON SOME SIMPLEST PROCESSES OF ORGANIZATION OF EYE AND BRAIN RUDIMENTS IN URODELAN EMBRYOS

By G. V. LOPASHOV

(Communicated by A. A. Zavarzin, Member of the Academy, 28.I.1945)

When carrying out experiments of transplantation of organ rudiments between different embryos, one often fails to discriminate between the results which arise autonomously and those dependent upon certain factors of the whole host embryo. And it is quite natural, for several agents belonging to the integer whole embryo will necessarily be present in the new place to which the rudiment was grafted. But by means of the method of embryonic explantation([5]) one succeeds, however, to segregate some elementary processes leading to the organization of rudiments. Outside the organism we become able to determine the agents which are present in whole embryos and which are responsible for the integrity of the embryos themselves and of their separate parts. The present paper is an attempt to expose certain data obtained with the aid of this method, bearing upon the appearance of normal organization and cellular structure of eye and brain rudiments in Urodelan embryos.

Explantation was practised with: 1) the I, II, III and IV quarters of the neural plate or of the mesoderm by which it is underlain, together with fragments of the belly ectoderm; 2) the same portion of the neural plate + the underlying mesoderm of the same or of another quarter of the neural plate + the abdominal ectoderm. The parts of the neural plate were isolated either together with the neural fold, or without it. Both neural plate and mesoderm were taken from the stage of early neurula; the ectoderm, from the stage of early gastrula up to the middle neurula. As material for these experiments served *Triturus taeniatus*, *Tr. vittatus* and *Amblystoma mexicanum*. The explant lived 2—16, mostly during 9—12 days. A total of over 1000 explantations were performed. The experiments were carried out during 1937—1940.

Types of the Development of Explants. According to the course of development and to structure, the explants can be classed into three groups, each of which offers a different picture of organ formation.

1. Irrespective of whether the ectoderm gives or gives not rise to the induction of brain, the explant is found to be entirely surrounded by the ectoderm, in which it is found freely to float. In the course of their development these types of explants are subject to different degrees of swelling.

2. Similar explants, settled at the bottom of the dish. Part of their ectoderm proliferates over the bottom in the form of a membrane. The ectoderm of these explants is always tight, and they remain compact throughout their development period.

3. The explanted (or induced) brain tissue lies at the surface of the explant which is partly or entirely covered with it. Explants of this type do usually also preserve solid structure.

Structure of Eye under Varying Developmental Conditions. Within the second group of the explants the eyes are invariably normal structures. This is true both of eyes that developed from the I section of the neural plate, and of eyes that were induced in the ectoderm by brain or mesoderm, or by the mesoderm of the I section in the neural plate of the II section. They consist of stratified retina, tightly invested with tapetum, and are cup-like in shape. They induce normal lenses. The eye structure of explants within the first group is dependent upon the space they are allowed to occupy in the course of their development. A number of variations in the eye shape are observed, *viz.*: a) relatively normal eyes, with a retina tightly invested with tapetum (largely in more compact explants); b) eyes representing vesicles of tapetum; only a portion of the wall of these vesicles is transformed into 1, 2, less frequently into 3 or 4 retinae, which are the smaller in size, the more extended is the explant; the complexity of the retinal structure also develops with size; c) eyes consisting entirely of big vesicles of tapetum; this type of eyes is encountered in atrongly extended explants only. Two types of eye development are observed within the third group. If developed inside the brain, the retina is a normal structure. But the tapetum which surrounds it under normal conditions is entirely lacking in this case. The retina is surrounded with brain, and at its surface only isolated pigmented cells are observable, which are entirely absent in older explants. In such cases where the eyes are protruding, their surface being directly bathed with the liquid medium, they are the shape of compact vesicles consisting of highly cylindric cells. Such type of «eyes» has already been described by Holtfreter ([7]).

Structure of Brain under Different Developmental Conditions. Similar relations are also found to exist in the development of brain. If developed in explants of the second group, the brain preserves its typical structure: thick lateral walls (giving rise to neuroblasts, glia and ependyme), and a thin roof, consisting of flat epithelium-like cells, of an ependymal type. The explants of the first group display all transitional stages from brains of this type to vesicular formations made up of epithelium-like cells, whose walls give room to the formation of individual cell accumulations, showing the structure of fragments of brain walls. The number of these accumulations varies from 1 to 2 in the posterior sections of the brain: they correspond to those portions of the brain from which its walls would develop under normal conditions. Unlike this, in the anterior sections of the brain numerous protrusions are often observed, fragments of the wall type being formed at their distal apices. The structure of these protrusions is much similar to that of the vesicular eyes bearing a small retina in their distal part. The protrusions are especially abundant within the region of the epiphysis and eyes, where the structures seem to imitate that of the eye. Sometimes several epiphyses are found to lie one behind another. In other cases fragments of the wall-type are encountered both at the bottom and the roof of the common vesicle, so that the cross-section of the brain assumes rather a radial structure. If the brain has been explanted together with the underlying mesoderm, it always preserves its typical structure. Larger fragments do more closely approach their normal structure than smaller ones. In explants belonging to the 3rd group the brain which lies at the surface is always quite compact, its structure resembling that of the wall of the brain.

In the case of purely cerebral explants, consisting of pure brain tissue, the fibrous substance is located deep within the explant, the polarity of the brain cells being a radial one. If the brain is in contact with the mesoderm which lies either within or outside it, the fibrous substance is turned towards the mesoderm, and the cells are polarized from the mesoderm to the liquid medium ([6, 7]). The same type of polarization is observed in the case the brain is in contact with the compact mass of the ectoderm.

Connexion between Brain and Eye. In the case where brain and eye develop largely into an epithelium-like tissue, it is of interest to observe transition into the latter. In the walls of the single elongated vesicle, the

cells of the «roof» of the brain are not infrequently substituted for by cells of the tapetum, which they resemble greatly in everything save for their colour. In one case the eye was found to be connected with the brain by a narrow tube consisting of epithelium-like pigmented cells reaching the very wall of the brain. This was evidently the optic nerve which, like the brain or the eye, had developed into an epithelium-like structure instead of a compact one.

Discussion. 1. By comparing such cases where the brain and the eye develop in a way approaching the normal one, with such cases where they develop as vesicles (some of them showing swellings at unusual places), it is possible to determine the fundamental distinctive features of either of these types of development. In case the space is limited because of the pressure produced by the external ectoderm, there is accumulation of eye and brain cells; in case there is no such pressure, there is no accumulation of cells either, or else it is less conspicuous, both in normal and atypical loci. Within the zones of accumulation, there is observed a development of cells of brain wall and of retina, whereas within such zones as show no accumulation, there is a development of epithelium-like cells.

2. This should be taken to mean that at the stage of neurula the cells of the neural plate may give rise to any kind of cells of the brain walls, of its roof, of any retina-cells (rods, cones, glial elements, neuroblasts), cells of the tapetum, cells of the optic nerve. If the segregation of eye-cells from the general cells of the neural plate seems to proceed under a special influence of the anterior section of the mesoderm([1],[2],[9]), the isolation of the retina, in the eye, and of the walls, in the brain, must evidently be due to the fact that at a definite locus the cells of these rudiments form accumulations, whereas in others they stretch as a single layer.

3. The accumulation of eye and brain cells both at normal and atypical loci shows that the tendency to cell accumulation is not fully predetermined and localized at the time of their appearance. The fact that the accumulation of cells may occasionally be lacking goes also to support this assumption. The cells may possibly gather into accumulations all over the surface of the rudiments, and are localized owing to some external influences. To these belong, in the first place, the condensing action of the epidermis of intact embryos, and, secondly, the formation of contact zones. The inducing effect of the mesoderm consists in that it is responsible for the formation of qualitatively specific cells of the «brain as such», or of the «eye as such»; and it is not until the next developmental stage that the mesoderm begins to play, as it seems, an essentially localizing part in the development of the brain. At the contact zone with the axial mesoderm the cells of the brain adhere to the latter, which results in their accumulation in this zone; the process of accumulation of brain cells thus started stimulates in its turn the condensation of cells of the axial mesoderm. This plays an important rôle in the development of the latter, too, which was duly emphasized in Yamada's experiments ([10]). The eye-cells are accumulated at the locus of contact with epidermis and lens ([3]); the latter is localizing the process of their condensation, the cells of the retina do likewise contribute to the local accumulation of the lens-cells, which does by no means preclude any other kind of interaction between the eye and the lens. At such loci as correspond to normal places of cell accumulation, the latter is due to the condensing effect of the contact zone already in operation. Owing to its higher density, the accumulation of cells thus started may induce at its place the formation of ever new cells, and play in this way an organizing rôle in their further development.

4. Unlike the induction of brain and eye, a mere accumulation of their cells may suffice to determine the subsequent phases of the histological future of their derivatives. This may possibly be due to the fact that certain substances produced by the condensing cells themselves may accumulate in a sufficient concentration within this zone. It is these products which may serve as agents determining the developmental trend of more differentiated cell accumulations, which they assume in contradistinction to the other cells of the initial rudiments. The varying concentrations of such substances may essentially contribute to the subdivision of the

bulk of the cell of an accumulation into different cell types. The endogenous origin of similar differentiation agents, with the simple modes of their distribution, constitutes probably one of the causes underlying the ability for self-organization and regulation, with which certain rudiments are highly endowed. These assumptions can to a certain degree be turned to good account when interpreting the rôle played by the volume of rudiments in their differentiation ([4],[8]).

5. The above hypothesis as to the relations existing between the development of eye and brain is likely to be helpful in the attempts to inquire into the nature of many a case of «self-differentiation». If a group of homogeneous cells undergoes a number of stages of segregation and differentiation due to no visible specific influence produced by other parts of the embryo, one cannot help thinking that these developmental stages had been predetermined for them at the time of their induction. But if their destiny is dependent on the influence of their own products, this interpretation should certainly be discarded. In this case, however, the effect is also dependent on the concentration of the substances produced by the cells, and this concentration is for each cell controlled by the relations between the cell and its surroundings, especially by the size and structure of the cell accumulations and the conditions under which these are formed. If we are to consider the formation of such products as a function of the cell as a whole, then, in so far as they are likely to be produced by any cell in the same way, the presence of these products in all cells would determine the individual destiny of each of them in a similar manner. We thus are brought to assume that the formation of such products is the function of one of the cell components, while their influence is applied to another component; the process of determination of the cells in such rudiments as the brain or the eye is probably of the nature of an interaction between the intercellular components, which is, however, brought about through the mediation of the total mass of the cells composing the rudiment and its spatial structure.

Institute of Cytology, Histology and Embryology.
Academy of Sciences of the USSR.

Received
28. I. 1945.

REFERENCES

[1] H. B. Adelmann, J. exp. Zool., 57, 223 (1930). [2] A. L. Alderman, J. exp. Zool., 70, 205 (1935). [3] N. I. Dragomirov, Bull. Acad. Sci. USSR, sér. biol., 741 (1939). [4] D. P. Filatov, Zool. J. Physiol, 51, 589 (1932). [5] J. Holtfreter, Arch. Entw.-Mech., 124, 404 (1931). [6] J. Holtfreter, ibid., 127, 619 (1933). [7] J. Holtfreter, Arch. exp. Zellforsch., 23, 169 (1939). [8] G. V. Lopashov, Biol. Zbl., 55, 606(1935). [9] G. V. Lopashov, C. R. Acad. Sci. URSS, XV, 286 (1937). [10] T. Yamada, Fol. anat. japon., 18, 569 (1939).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVIII, № 8

PARASITOLOGY

EVOLUTION OF *EURYTOMA* SPECIES *(HYMENOPTERA, CHALCIDIDAE)* IN CONNEXION WITH THE GEOGRAPHICAL DISTRIBUTION OF THEIR FOOD-PLANTS OF THE SUBFAM. *PRUNOIDEAE*

By M. N. NIKOLSKAIA

(Communicated by E. N. Pavlovsky, Member of the Academy, 8.XI.1944)

Three closely related species of *Eurytoma* Ill. feed on stones of *Prunoideae*; these have in general a similar life history. The females oviposit into immature fruits, and the larvae develop feeding on kernels. The infested fruit usually drops when still unripe with the larva hibernating in it; pupation takes place in spring and the adult appears when the ovaries are present on the trees.

The first species, *Eurytoma amygdali* Enderlein, was described in Europe from stones of cultivated almond in 1907. In 1908 Schreiner reared the species of plums in the Astrakhan province, and described it by the name *Eurytoma schreineri* Mayr, which was reduced to a synonym of *E. amygdali* End. by Rodzianko in 1913.

This species is widely spread in the west Mediterranean (South of the W. Europe, Cyprus, Syria, Palestine) and the European part of the USSR, up to Kursk region in the North. In the Mediterranean it was reared from cultivated almond, *Amygdalus communis* L., being such a severe pest of its fruit that a control was prescribed by legislation. According to data of Russian entomologists, *E. amygdali* End. infests the fruit of different varieties of *Prunus domestica* L. *and Armeniaca vulgaris* Lam., but is also found in *Prunus spinosa* L. and *P. divaricata* Ldb.

The other species, *Eurytoma samsonovi* Vas., was reared from apricots in Ferghana, and described by I. V. Vasilev in 1915. It is common all over Central Asia (Kostina, 1936). In 1932 Fletcher recorded it among the apricot pests from north India, while Pruthi and Batra (1939) consider it to be of wide distribution in India on apricot, plum and peach.

A new species, *Eurytoma maslovskii*, was reared by the late entomologist N. N. Maslovsky from plum and apricot fruit in the Ussurian region, during the years 1932—1934. The description he gives is as follows.

Eurytoma maslovskii sp. nov. F e m a l e. Length 7—8 mm. Body black, apical parts of femora, all tibiae slightly darkened in the middle, and tarsi as well as ovipositor tip, dark yellow. Thimble-like shallow punctures on head and thorax close to each other, so that the interspace between them is narrow, and white pubescence seems to be dense. Propodeum wrinkled-reticulate, with a longitudinal groove not very distinctly separated from side reticulation and with an indistinct median carina. Petiole small, scarcely seen. Abdomen long, ovate, pointed at apex. Front wings fuscous, with marginal vein longer than postmarginal one. Ring-joint of antennae narrow, first funicular joint 2.5 times as long as wide.

M a l e. Length 6—7 mm. Similar to female. Abdomen short, petiole slightly longer than hind coxae, gradually increased towards abdomen. Antennae with

joints 2—4 distinctly pedunculate at apices, the narrowed part being as wide as long.

On plums and apricots in Ussurian region, Far East. To distinguish the three species the following key is suggested.

Key of Species. 1 (2) Head and thorax black. Abdomen reddish-brown. Scape and legs except coxae dark-yellow. Body length in female 5.8—6.4 mm, male 4—4.2 mm. Apricots, plums and peaches in Central Asia and India.

1. *Eurytoma samsonovi* Vas. 2 (1) Body black, only tips of femora, tibiae and tarsi yellowish. 3 (4) Tibiae dark, with yellow proximal and apical ends. Thorax more scarcely punctuated. Front wings nearly hyaline with a fuscous spot under marginal vein, which is slightly shorter than postmarginal one. Propodeum with a deep, well-marked groove in the middle and distinct longitudinal carina. Abdominal petiole in female conspicuous, with three spines on dorsal surface. Male petiole as long as hind coxae, of equal width throughout its length. Female antennae with a ring-joint rather broad and first funicular joint 2 times as long as wide. Joints 2—4 of male antennae indistinctly pedunculate at apices. Body length in female 4—6 mm, in male 6—7 mm.

Amygdalus communis L., *Prunus domestica* L., *P. spinosa* L., *P. divaricata* Ldb. and *Armeniaca vulgaris* Lam.

South of Europe, Cyprus, Syria, Palestine, European part of the USSR up to Kursk region in the North.

2. *Eurytoma amygdali* End. 4 (3) Tibiae dark yellow, slightly blackened in the middle. Thorax closely punctate. Front wings fuscous. Marginal vein a little longer than postmarginal one. Propodeum with less distinct median groove and interrupted longitudinal carina. Female petiole inconspicuous. Male petiole somewhat longer than hind coxae, gradually increasing towards abdomen. Female antennae with a narrow ring-joint and first funicular joint 2.5 times as long as wide. Joints 2—4 of male antennae distinctly pedunculate at apices. Body length in female 7—8 mm, in male 6—7 mm.

On plums and apricots in Ussurian region, Far East. Probably it will be found in N. China, Manshuria and Japan.

According to Kovalev and Kostina (1941) geographical distribution of some species subfam. *Prunoideae* is as follows:

Prunus spinosa L. grows chiefly in the forest and steppe zone of W. Europe from Atlantic Ocean to S. Scandinavia, Mediterranean coast, Asia Minor, to N.-W. Iran, the Caucasus, European part of the USSR from the Crimea to the sources of the Dnieper and Volga, eastward to the source of the Tobol in W. Siberia.

P. domestica L. with all its varieties is known only in culture and supposed to originate by hybridization of *P. spinosa* L. with *P. divaricata* Ldb., which is cultivated as well (*P. cerasifera* Ehrh.).

Amygdalus communis L. grows in wild condition on the rocky slopes (800—1600 m above the sea level) in S. Transcaucasia, mountains of Turcomania, Tian-Shan, Iran, Armenian Kurdistan, E. Asia Minor, closely coinciding in distribution with *P. divaricata* Ldb. Cultivated *A. communis* L. is widely spread all over the dry subtropics.

Armeniaca vulgaris Lam. is the only species of the genus growing wild in Central Asia (500—1200 m above sea level), and giving many a variety in culture of S.-W. Asia. Another species, *A. dasycarpa* Pers., is known only in culture of Anterior and Central Asia, seeming to be a hybrid of *A. vulgaris* Lam. and *P. cerasifera* Ehrh.

The area of distribution of the remaining six *Armeniaca* species extends from the S. Ussurian region and N. Korea through Manshuria and N. China to the N.-E. Mongolia.

Each of the three species of *Eurytoma* Ill. has a definite area of distribution that may be compared with the corresponding area of its food-plant. Thus, *Prunus spinosa* L. is considered to be the original food-plant along with the extension of whose geographical area the evolution of *E. amygdali* End. took place; its connexion with *P. domestica* L., *P. spinosa* L., *P. divaricata* Lbd. should have

arisen later. The same conclusion is quite natural with respect to the cultivated almond.

Evolution and distribution of *Eurytoma samsonovi* Vas. went along that of *Armeniaca vulgaris* Lam., from which it formed a transition to different varieties of apricot, plum and peach.

The most probable original food-plant of *Eurytoma maslovskii* Nik. was *Armeniaca manshurica* (Koene) Skwortz., distributed from the lake Hanka and S. Ussurian region throughout E. Manshuria to the woods of N. Korea, and cultivated in S. Ussurian region and N. Manshuria. Its connexion with N. China species: *A. ansu* (Komar.) Kost. and *A. mume* Sieb., seems to be possible, the species being grown in Japan, Korea and E. China and giving a lot of apricot varieties of S.-E. Asia.

Similar considerations are found in a paper by Tranzschel and Litvinov (1939) on the fungi of the genus *Tranzschelia* Arth. Several species are distinguished in the former *T. pruni-spinosae* Pers. with ecidiae on anemones and uredo- and theleytospores on *Prunoideae*. The type species of *T. pruni-spinosae* (Pers.) is reported to be the form with uredo- and theleyto-spores on black thorn and plum distributed all over Europe. A species from Japan (*T. japonica* Tranz. and Litv.) was detected on apricots *A. mume* Sieb. and *A. ansu* (Komar.) Kost. A new species *T. microcerasi* Tranz. and Litv. was collected on species of the genus *Cerasus* Jurs. (*Microcerasus* Webb.) in Central Asia.

Zoological Institute.
Academy of Sciences of the USSR.

Received
8. IV. 1944.

REFERENCES

[1] И. В. Васильев, Тр. Бюро энт., XI, 7 (1915). [2] К. И. Водинская, Защ. раст., 1, 97 (1932). [3] Н. В. Ковалев и К. Ф. Костина, Флора СССР, X, 1941. [5] К. Ф. Костина, Бюлл. пр. бот., 83 (1936). [5] М. Н. Никольская, Русск. энт. обозр., 25, 1—2, 121 (1933). [6] М. Н. Никольская, Инф. бюлл. карант. раст., 6, 6 (1939). [7] Е. В. Пузанова-Малышева, Русск. энт. обозр., 24, 3—4, 166 (1930). [8] В. Н. Родзянко, О миндальном семяеде, повреждающем сливы и абрикосы в Астраханской губ., Астрахань, 1913. [9] В. Г. Траншель и М. А. Литвинов, Бот. журн., 24, 3, 47 (1939). [10] А. Устинов, Вісн. Харків. С.-г. Інст. (1925). [11] P. Chobardiev, Rapp. Ann. St. Agr. Sofia (1927). [12] J. Aharoni, Bull. Agr. Intell., 7, 11 (1916). [13] W. Bevan, Ann. Rep. director Agr. Cyprus for 1917—1918, Nicosia (1918). [14] G. Enderlein, Z. Hym. Dipt., 7, 303 (1907). [15] T. B. Fletcher, Sci. Rep. Imp. Inst. Agr. Res. Pusa (1932). [17] G. Lesne, P. Ann. Serv. Epiph., VI (1919). [17] W. J. Phillips, J. Agr. Res., 34, 8 (1927). [18] H. S. Pruthi and R. N. Batra, Ind. J. Agr. Sci., 9, 2, 277 (1930). [19] I. T. Schreiner, Z. wiss. Insekt. Biol., 4, 1—2 (1908).